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ON THE ANALOGY BETWEEN MATHEMATICAL PROBLEMS  
OF NON-LINEAR FILTERING AND QUANTUM PHYSICS<sup>(1)</sup>

by

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20. ABSTRACT (Continue on reverse side if necessary and identify by block number) -The main thesis of this paper is that there are striking similarities between the mathematical problems of stochastic system theory, notably linear and non-linear filtering theory, and mathematical developments underlying quantum mechanics and quantum field theory. Thus the mathematical developments of the past thirty years in functional analysis, lie groups and lie algebras, group representations and probabilistic methods of quantum theory can serve as a guide and indicator to search for an appropriate theory of stochastic systems. In the current state of development of linear and non-linear filtering theory, it is best to proceed			

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## 20. Abstract cont.

by "analogy" and with care, since "unitarity" which plays such an important part in quantum mechanics and quantum field theory is not necessarily relevant to linear and non-linear filtering theory. The partial differential equations that arise in quantum theory are generally wave equations, whereas the partial differential equations arising in filtering theory are stochastic parabolic equations. Nevertheless the possibility of passing to a wave equation by appropriate analytic continuation from the parabolic equation, reminiscent of the current program in euclidean field theory, should not be overlooked.

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## 1. Introduction.

The main thesis of this paper is that there are striking similarities between the mathematical problems of stochastic system theory, notably linear and non-linear filtering theory, and mathematical developments underlying quantum mechanics and quantum field theory. Thus the mathematical developments of the past thirty years in functional analysis, lie groups and lie algebras, group representations and probabilistic methods of quantum theory can serve as a guide and indicator to search for an appropriate theory of stochastic systems. In the current state of development of linear and non-linear filtering theory, it is best to proceed by "analogy" and with care, since "unitarity" which plays such an important part in quantum mechanics and quantum field theory is not necessarily relevant to linear and non-linear filtering theory. The partial differential equations that arise in quantum theory are generally wave equations, whereas the partial differential equations arising in filtering theory are stochastic parabolic equations. Nevertheless the possibility of passing to a wave equation by appropriate analytic continuation from the parabolic equation, reminiscent of the current program in euclidean field theory, should not be overlooked.

To develop these ideas, it is best to begin with a reasonably general non-linear filtering problem:

Let  $(\Omega, \mathcal{A}, P)$  be the underlying probability space and let  $x_t$  denote a scalar-valued diffusion process which is the solution of the following Ito stochastic differential equation:

$$(1.1) \quad dx_t = f(x_t)dt + g(x_t)dw_t, \quad$$

where  $w_t$  is standard Brownian motion and  $f$  satisfies appropriate assumptions so that

(1.1) has a unique solution in the sense of Ito. We shall refer to (1.1) as the physical process. Let

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(1) There is no difficulty in generalizing to the vector case.

(1.2)  $z_t = h(x_t)$  with  $z_t \in L^2(\Omega, \mathcal{A}, P)$ , denote the signal process, and we observe the signal in the presence of another Wiener process

$$(1.3) \quad y_t = \int_0^t z_s ds + \eta_t. \quad \text{We shall refer to (1.3) as the observation equation.}$$

We make the assumption that  $(x_t, \eta_t)$  are independent. Let  $F_t^y$  denote the  $\sigma$ -field generated by  $\{y_s | 0 \leq s \leq t\}$ . The problem of non-linear filtering is to recursively compute

$$(1.4) \quad E[\phi(x_t) | F_t^y], \quad \text{where } \phi \text{ say is a bounded, continuous function. } E[\phi(\cdot) | F_t^y]$$

denotes conditional expectation with respect to  $F_t^y$ . The solution to this problem can be obtained by Functional Integration and the Cameron-Martin-Gisanov formula.

Define a new measure  $\tilde{P}$  on  $(\Omega, \mathcal{A})$  by the Cameron-Martin-Gisanov transformation:

$$(1.5) \quad \frac{dP}{d\tilde{P}} = \exp \left( \int_0^t z_s dy_s - \frac{1}{2} \int_0^t z_s^2 ds \right).$$

Under this new measure, the probability distribution of  $x_t$  remains invariant, but  $y_t$  and  $x_t$  are independent and  $y_t$  is standard Brownian motion.

Let  $\pi_t$  denote the conditional expectation operator. Then a standard application of the theory of conditional expectations gives us:

$$(1.6) \quad \pi_t(\phi) = E[\phi(x_t) | F_t^y] = \frac{E[\phi(x_t) \frac{dP}{d\tilde{P}} | F_t^y]}{\tilde{E}[\frac{dP}{d\tilde{P}} | F_t^y]}$$

where  $\tilde{E}$  denotes expectation with respect to the  $\tilde{P}$ -measure. The mapping  $\phi \mapsto \pi_t(\phi)$  is defined to be the filter for the stochastic system (1.1)-(1.3).  $\pi_t$  itself can be thought of as a measure-valued stochastic process.

For what follows it is convenient to rewrite (1.6) in the form of an input-state-output relation. For this purpose define

$$(1.7) \quad \rho_t : \phi \mapsto \tilde{E}[\phi(x_t) \frac{dP}{d\tilde{P}} | F_t^x].$$

Then (1.6) may be rewritten as

$$(1.8) \quad \pi_t(\phi) = \frac{\rho_t(\phi)}{\rho_t(1)}, \text{ where}$$

1 denotes the constant function 1 for all x.

In the above  $\rho_t$  is to be thought of as the state of the filter and equation (1.8) as the state-output relation of the filter.

It is instructive to view (1.7) and (1.8) in the light of Gelfand-Naimark-Segal construction of states and representations. The functions  $x_t \mapsto \phi(x_t)$  are the observables of the physical system (1.1). The formula (1.8) computes the conditional statistics of the observables  $\phi$  given the observation program  $\{y_s | 0 \leq s \leq t\}$ . The analogy with the algebraic theory of quantum mechanics is striking, the notable difference being that the idea of computing conditional statistics based on an observation program seems to be absent in physics. This viewpoint turns out to be important in the definition of generalised observables for quantum systems as probability operator-valued measures.

It can be shown that  $\rho_t$  has density  $q_t$  which satisfies a stochastic partial differential equation

$$(1.9) \quad dq_t = L_0^* q_t dt + L_1 q_t dy_t, \quad ,$$

where  $L_0^*$  is the formal adjoint of the diffusion process generator of (1.1) and  $L_1$  is the operator : multiplication by  $h(x)$ . This is the <sup>Duncan-</sup>Mortensen-Zakai equation and is the fundamental equation of non-linear filtering. The density  $q_t(z, y_0^t)$  has a representation as a function space integral

$$(1.10) \quad q_t(z, y_0^t) = \int_X \exp(\int_0^t h(x_s) dy_s - \frac{1}{2} \int_0^t h^2(x_s) ds) d\mu^z(x)$$

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where the integration is over the path space of  $x$  with  $x_0 = 0$  and  $x_t = z$ .<sup>(1)</sup>

Now in equation (1.9) the observation  $\{y_s | 0 \leq s \leq t\}$  is given to us, and if we rewrite equation (1.9) in the following suggestive form using Stratanovich calculus:

$$(1.11) \quad \frac{dq_t}{dt} = (L_0^* - \frac{1}{2} L_1^2) q_t + \dot{y}_t L_1 q_t \quad ,$$

then equation (1.11) is the analogue of a euclidean (imaginary time) quantum field a time-dependent with  $\wedge$  external force defined by the observations. We say that the analogy is to euclidean quantum field (as opposed to euclidean quantum mechanics) since  $q_t$  is a measure-valued stochastic process. This paper is concerned with a systematic investigation of this point of view. In particular, in this picture the Kalman filter occupies the role of the free quantum field.

This paper is divided into six sections. In section 2 we discuss the relationship between Dirichlet and Schrödinger operators and show that it is possible to associate a stochastic process with the ground state measure of Schrödinger operators. Section 3 is devoted to the Bayes formula on non-linear filtering and describes the Feynman point of view for non-linear filtering. Section 4 is concerned with the construction of Fock space and discussing its role in non-linear filtering. In section 5 we argue that the Lie algebra of operators  $L = LA[L_0^* - \frac{1}{2} L_1^2, L_1]$  has an important role to play in non-linear filtering. In section 6 we discuss the question of representation of the filter.

This is a semi-expository paper and we have tried to concentrate on the ideas involved and emphasize a certain point of view. The ideas come from constructive quantum field theory as emphasized by Nelson and Segal, recent developments in system theory and the theory of group representations. It is our hope that this paper will go towards pointing in a small way the conceptual depth of stochastic system theory which is still in its infancy.

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(1) That is an integration over the path space of  $x$ , conditioned by  $x_0=0$  and  $x_t=z$ .



## 2. The Feynman-Kac Formula, Dirichlet and Schrodinger Operators.

### 2.1 Introduction

Recent work on non-linear filtering theory, euclidean quantum field theory and the stochastic mechanics of Nelson make extensive use of the Feynman-Kac formula and the interplay between Schrodinger and Dirichlet operators. In particular, a stochastic process associated with the ground-state measure of Schrodinger operators turns out to be important. In this section we give an account of these ideas and the related theory of hypercontractive semigroups.

### 2.2. Preliminaries.

Throughout this paper integration with respect to Wiener measure, the Brownian bridge measure and the oscillator measure will be important. The relation of Wiener measure to the Laplacian and the semigroup generated by the Laplacian will also be important. For simplicity we shall be concerned with scalar-valued stochastic processes. There is no difficulty in generalising these ideas to vector-valued stochastic processes.

Let  $\Omega = C(\mathbb{R}_+; \mathbb{R})$  and let  $W_t: \omega \rightarrow W_t(\omega) = \omega_t^{(*)}: \Omega \rightarrow \mathbb{R}$  be the  $t^{\text{th}}$  coordinate function. We denote by  $\tau_t$  the right shift on  $\Omega$ . If we denote by  $\mathcal{F}_t$  the smallest  $\sigma$ -field with respect to which  $\{W_s \mid 0 \leq s \leq t\}$  are measurable and by  $\mathcal{F} = \sigma\{W_t \mid t \geq 0\}$ , then  $\mathcal{F}$  is the Borel  $\sigma$ -field of  $\Omega$  and on  $(\Omega, \mathcal{F})$  there is a unique probability measure  $W$  such that  $W(W_0=0)=1$ , the random variables  $W_{t_1}-W_{t_0}, \dots, W_{t_n}-W_{t_{n-1}}$  are independent, Gaussian, zero mean, with variance  $t_j - t_{j-1}$  for the  $j^{\text{th}}$  increment. This measure is Wiener-measure. For each  $x \in \mathbb{R}$ , the probability measure  $W_x$  is defined by path translation:

---

(\*)We shall also use the notation  $x(t)$  for  $x_t$  when dealing with processes.

$$\begin{aligned} W_x(B) &= W(\theta_x(B)) \quad B \in \mathcal{F}, \text{ where} \\ \theta_x(\omega)_t &= x + \omega_t, \quad x \in \mathbb{R}, \quad \omega \in \Omega, \quad t \geq 0. \end{aligned}$$

Alternatively one could have started with a Gaussian process  $\{W_t \mid t \geq 0, W_0 = 0 \text{ a.s.}\}$  which is zero-mean and has variance  $|t-s|$  and stationary independent increments and constructed Wiener measure as a path space measure on  $C(\mathbb{R}_+; \mathbb{R})$ . For a measurable real-valued function  $f$ , define

$$(2.1) \quad (P_t f)(x) = E_{W_x}(f(W_t))$$

where  $E_{W_x}$  denotes expectation with respect to  $W_x$ . If the right hand side makes sense,

$$(2.2) \quad (P_t f)(x) = \int_{\mathbb{R}} f(y) p(t, y; 0, x)$$

where  $p(t, y; 0, x)$  is the transition density of Brownian motion

$$(2.3) \quad p(t, y; 0, x) = \frac{1}{(2\pi t)^{\frac{1}{2}}} \exp\left(-\frac{(x-y)^2}{2t}\right), \quad t > 0, \quad x, y \in \mathbb{R}.$$

It is known that for  $p \in [1, \infty]$ ,  $(P_t \mid t \geq 0)$  is a strongly continuous contraction semigroup on  $L^p(\mathbb{R})$ . Its infinitesimal generator is  $-\frac{1}{2}\Delta$ , where  $\Delta$  is the Laplacian and  $\mathcal{D}(A) = H^p(\mathbb{R}) = \{f \in L^p \mid Df \in L^p \text{ in the sense of distributions}\}$ .

For certain applications we shall need to do integration with respect to conditional Wiener measure. From the properties of Wiener measure and the corresponding transition density of Brownian motion, for  $0 < t_1 < \dots < t_m$

$$W_x \{ \omega \mid \omega(t_1) \in A_1, \dots, \omega(t_m) \in A_m \} = \int_{\mathbb{R}} \prod_{i=1}^m p(x_{i-1}, x_i; t_i - t_{i-1}) \chi_{A_i}(x_i) dx_i$$

where  $\chi_A$  is the characteristic function of  $A$ ,  $x_0 = x$ ,  $t_0 = 0$ . We can then construct a measure  $W_{x, y; t}$  on the continuous paths on  $[0, t]$  with  $W_0 = x$ ,  $W_t = y$

with probability 1,  $0 < t_1 < \dots < t_m < t$

$$\begin{aligned} W_{x,y;t} \{ \omega | \omega(t_1) \in A_1, \dots, \omega(t_m) \in A_m \} = \\ = \int_{\mathbb{R}} \prod_{i=1}^m p(x_{i-1}, x_i; t_i - t_{i-1}) \chi_{A_i}(x_i) p(x_m, y; t - t_m) dx_i \end{aligned}$$

$W_{x,y;t}$  is called conditional Wiener measure. We have

$$(2.4) \quad \int f(\omega) dW_x = \int dy \left( \int f(\omega) dW_{x,y;t} \right)$$

for  $f$  which are functions of the values the path takes on  $[0, t]$ .

This measure corresponds to the Brownian bridge process, which is the Gaussian process  $\beta_s$ , say on  $[0, 1]$  with covariance

$$E(\beta_s \beta_t) = s(1-t), \quad 0 \leq s \leq t \leq 1$$

In terms of this process we can write (2.4) as

$$(2.5) \quad \int f(\omega_s) dW_0 = \int f\left(\left(1 - \frac{s}{t}\right)x + \frac{s}{t}y + \sqrt{\frac{s}{t}} \beta_s\right) p(x, y; t) dx dy d\beta$$

where  $d\beta$  represents the measure on the path space of  $\beta$ .

The two processes which we have dealt with are not stationary Gauss-Markov processes. For this reason it is often important to deal with the oscillator process which is the family of Gaussian random variables

$\{q_t, -\infty < t < \infty\}$  with covariance  $\frac{1}{2} \exp(-|t-s|)$ . It is related to the operator  $L_0 = -\frac{1}{2} \frac{d^2}{dx^2} + \frac{1}{2} x^2 - \frac{1}{2}$ , the Harmonic Oscillator Hamiltonian.

### 2.3 The Feynman-Kac Formula.

Both in filtering theory and in quantum physics we are required to deal with an operator on  $L^2(\mathbb{R})$  (say),

$$H = -\frac{1}{2}\Delta + V(x)$$

where  $\Delta$  is the Laplacian and we need to compute  $\exp(-tH)$ , the semi-group generated by  $-H$ . The Feynman-Kac formula provides us with a representation of the operator  $T_t = \exp(-tH)$  as a Wiener Integral. The first approach exemplified in the work of Nelson uses the Trotter Product formula to make sense of the formula but requires information on the self-adjointness of the operator  $H$ . The second approach proceeds by writing the formula first, proving that it represents a strongly continuous one-parameter semigroup and then its unique infinitesimal generator is computed. The second approach uses probabilistic techniques and turn out to be the more general one.

If the potential is bounded above then a general form of the Feynman-Kac formula can be obtained using martingale methods. For this purpose and for later use we introduce certain probabilistic machinery.

Let  $(\Omega, F, P)$  be a probability space and  $(F_t, t \geq 0)$  a non-decreasing family of sub  $\sigma$ -fields of  $F$ . Let  $s \geq 0$  be arbitrary and  $a: [s, \infty[ \times \Omega \rightarrow \mathbb{R}$  and  $b: [s, \infty[ \times \Omega \rightarrow \mathbb{R}$  be bounded progressively measurable functions. For any  $f \in C^2(\mathbb{R})$ , define  $A_t f$  for  $t \geq s$  by

$$(2.6) \quad (A_t(q)f)(x) = \frac{1}{2} a(t, q) \frac{\partial^2 f}{\partial x^2}(x) + b(t, q) \frac{\partial f}{\partial x}(x)$$

If  $\xi(\cdot, \cdot)$  is any progressively measurable function from  $[s, \infty[ \times \Omega \rightarrow \mathbb{R}$ , then  $(A_t(q)f)(\xi(t, q))$  defines another progressively measurable function of  $t$  and  $q$ . Then it can be shown that for any  $\xi$ ,  $a$  and  $b$  satisfying the above,

$$(2.7) \quad f(t, \xi(t)) - \int_s^t \left( \frac{\partial}{\partial u} + A_u \right) f(u, \xi(u)) du$$

is a martingale relative to  $(\Omega, \mathcal{F}_t, P)$  for all  $t \geq s$ ,  $\forall f \in C^{1,2}([0, \infty[ \times \mathbb{R})$ .  
 A process  $\xi$  satisfying (2.7) will be referred to as an Ito process with drift  $b$  and covariance  $a$ . The Feynman-Kac formula depends on the following observation.

Let  $\xi$  be an Ito process relative to  $(\Omega, \mathcal{F}_t, P)$  with drift  $b$  and covariance  $a$ . Then for any progressively measurable function  $V: [s, \infty[ \times \Omega \rightarrow \mathbb{R}$  which is bounded below and  $f \in C^{1,2}([0, \infty[ \times \mathbb{R})$

$$(2.8) \quad e^{\int_s^t -V(u) du} f(t, \xi(t)) - \int_s^t \left( \frac{\partial f}{\partial u} + A_u f - V(u) f \right) (u, \xi(u)) e^{\int_s^t -V(\sigma) d\sigma} du$$

is a martingale after time  $s$ . (1)

Now define the operator

$$(2.9) \quad T_t^V f(s, x) = \int e^{\int_s^t -V(u) du} f(t, \xi(t)) d\mu_{sx}(\xi)$$

where  $d\mu_{s,x}(\xi)$  denotes the measure on the path-space of  $\xi$  given  $\xi(s) = x$ .  
 Then we can show that  $T_t^V$  is a one-parameter strongly continuous semi-group and from (2.8) we conclude that its differential generator is  $A_u + V(u)$ .  
 (2.9) is the Feynman-Kac formula. When we specialize to the case when  $\xi(t)$  is Brownian motion, then  $A_u = \frac{1}{2} \frac{d^2}{dx^2}$  and the formula reads setting  $s=0$ .

$$(2.9') \quad (T_t^V f)(x) = E_{W_x} \left[ \exp \left( \int_0^t -V(W(s)) ds \right) f(W(t)) \right].$$

Now this formula turns out to be valid for a much larger-class of unbounded potentials and the semigroup  $T_t^V$  has strong regularity properties.

---

(1) This argument is an exercise in Stroock-Varadhan: Multi-dimensional Diffusion Processes, Springer-Verlag, N.Y., 1979.

The more general class of potentials we consider are: (H1)  $V$  is a measurable function on  $\mathbb{R}$  which can be written as  $V = V_1 - V_2$  with  $V_2 \geq 0$ ,  $V_2 \in L^p(\mathbb{R})$  for  $p > 1$  and  $V_1$  measurable, bounded below, such that for each compact set  $K$  in  $\mathbb{R}$ , there exists  $q(K) \geq 2$  s.t.

$$\int_K |V_2(x)|^q dx < \infty.$$

For this class of potentials, using the recent estimate (Berthier-Gaveau).

$$\forall t > 0, \forall r > 0, K(r,t) \triangleq \sup_{x \in \mathbb{R}} E_{W_x} \left( \exp(-r \int_0^t V(x_s) ds) \right) < \infty$$

and their refinements, the Feynman-Kac formula holds for the class of potentials governed by (H1). If we define the operator

$$T_t(f)(x) = E_{W_x} \left( \exp \left( - \int_0^t V(W_s) ds \right) f(W_t) \right),$$

we have

Theorem 2.1. (Carmona) For any  $q \in [1, \infty[$  and  $t > 0$ ,

- (i)  $T_t$  is a bounded operator on  $L^q$  and  $\|T_t\|_q \leq K(1,t)$
- (ii) If  $q'$  denotes the conjugate exponent of  $q$ , then for  $f \in L^q$ ,  $g \in L^{q'}$

$$\int_{\mathbb{R}} (T_t f)(x) g(x) dx = \int_{\mathbb{R}} f(x) (T_t g)(x) dx$$

- (iii)  $T_t$  is a strongly continuous semi-group on  $L^q$
- (iv) If  $\lim_{|x| \rightarrow \infty} V_1(x) = +\infty$ , then  $T_t$  is a compact operator on  $L^q$ . ▢

### 2.3. Schrodinger and Dirichlet Operators.

Let  $-H$  denote the infinitesimal generator of  $T_t$ . In quantum physics we often need to show that  $-H$  is the self-adjoint extension of  $-\frac{1}{2}\Delta + V$ , the imaginary-time Schrodinger operator. If  $V \geq 0$  this is well known. For the class of potentials given by (H1) if  $V \in L^q_{loc}$  then

Proposition 2.2. (Carmona)

$C_c^\infty(\mathbb{R}) \subset \mathcal{D}(H)$  and for  $f \in C_c^\infty(\mathbb{R})$ ,  $Hf = -\frac{1}{2}(\Delta + V)f$  where  $H$  is the infinitesimal generator of the semigroup  $T_t$  defined on  $L^q$ .  $\square$

In mathematical physics perturbed Hamiltonians are usually defined as sums of quadratic forms. We briefly review this.

Let  $H$  be a Hilbert space. A quadratic form is a map  $q: Q(q) \times Q(q) \rightarrow \mathbb{R}$ , where  $Q(q)$  is a dense linear subset of  $H$ , called the form domain, such that  $q(\cdot, \psi)$  is conjugate linear and  $q(\phi, \cdot)$  is linear for  $\phi, \psi \in Q(q)$ . If  $q(\phi, \psi) = \overline{q(\psi, \phi)}$  we say that  $q$  is symmetric. If  $q(\phi, \phi) \geq 0 \quad \forall \phi \in Q$ ,  $q$  is called positive and if  $q(\phi, \phi) \geq -M \|\phi\|^2$ , for some  $M$  we say  $q$  is semi-bounded.

Let  $A$  be a self-adjoint operator on  $H$ . By passing to the spectral representation of  $A$ ,  $A$  is multiplication by  $x$  on  $\bigoplus_{n=1}^{\infty} L^2(\mathbb{R}, \mu_n)$ . Let

$$Q(q) = \left\{ \psi_n(x) \mid \sum_{n=1}^N \left| \sum_{n=1}^N \int_{-\infty}^{\infty} x |\psi_n(x)|^2 d\mu_n < \infty \right. \right\}$$

for all  $\psi, \phi \in Q(q)$  define

$$q(\phi, \psi) = \sum_{n=1}^N \int_{-\infty}^{\infty} x \phi_n(x) \psi_n(x) d\mu_n$$

$q$  is called the quadratic form associated with  $A$  and writing  $Q(A) = Q(q)$ ,  $Q(A)$  is called the form domain of  $A$ .

Let  $q$  be a semi-bounded quadratic form,  $q(\psi, \psi) \geq -M \|\psi\|^2$ .  $q$  is called closed if  $Q(q)$  is complete under the norm

$$\|\psi\|_{+1} = \sqrt{q(\psi, \psi) + (M+1) \|\psi\|^2}.$$

If  $q$  is closed and  $D \subset Q(q)$  is dense in  $Q(q)$  in the  $\|\cdot\|_{+1}$  norm then  $D$  is called a form core for  $q$ . The following fact is important.

Theorem: If  $q$  is a closed semibounded quadratic form, then  $q$  is the quadratic form of a unique self-adjoint operator.  $\square$

We now define Schrodinger operators as form sums on  $L^2(\mathbb{R})$ . For  $f, g \in C_c^2(\mathbb{R})$  define

$$\varepsilon_0(f, g) = \frac{1}{2} \int_{\mathbb{R}} \nabla f(x) \overline{\nabla g(x)} dx$$

Integrating by parts

$\varepsilon_0(f, g) = (-\frac{1}{2} \nabla^2 f, g)_2$  and hence  $\varepsilon_0$  is closable. The form domain  $Q(\varepsilon_0)$  of  $\varepsilon_0$  is  $H^1(\mathbb{R})$ . Let  $V$  be a real-valued measurable fn. on  $\mathbb{R}$  and set

$$Q(V) = \{f \in L^2 \mid \int_{\mathbb{R}} |V(x)| |f(x)|^2 dx < +\infty\}$$

On  $Q(\varepsilon_0) \cap Q(V)$  define



$$\varepsilon(f,g) = \varepsilon_0(f,g) + (Vf,g) \text{ where}$$

$$(Vf,g) = ((\operatorname{sgn} V)|V|^{1/2}f, |V|^{1/2}g)$$

If we assume  $V = V_1 - V_2$  s.t.  $V_1$  bounded below and  $V_1 \in L^1_{loc}$ ,  $V_2 \in L^1$  then  $\varepsilon$  is the form of a unique bounded below self-adjoint operator  $H$  on  $L^2$  with form domain  $Q(H) = Q(\varepsilon_0) \cap Q(V)$ .

Let  $\mu$  be a Borel prob. measure on  $\mathbb{R}$  which satisfies

$$(2.10) \quad d\mu(x) = e^{-2h(x)} dx$$

where  $h$  is a real-valued, locally bounded, absolutely continuous function with first order partial derivatives in  $L^2_{loc}$ . For all  $f, g \in C^\infty_c(\mathbb{R})$ , define

$$(2.11) \quad \delta(f,g) = \frac{1}{2} \int_{\mathbb{R}} \nabla f(x) \overline{\nabla g(x)} d\mu(x)$$

Integrating by parts the right hand side, we get

$$(2.12) \quad \delta(f,g) = (Df,g)_\mu, \text{ where } (\cdot, \cdot)_\mu \text{ denotes the } L^2(\mu)\text{-inner}$$

product and

$$(2.13) \quad Df = -\frac{1}{2}\Delta f + \nabla h \nabla f$$

From (2.12)  $\delta$  is given by a symmetric operator and hence is closable.

Let  $\bar{\delta}$  denote the closure.  $\bar{\delta}$  is referred to as the Dirichlet form of  $\mu$  and  $D$  the associated Dirichlet operator.

We can prove that

Proposition 2.3. (Carmona) The form domain  $Q(\bar{\delta})$  is  $H^1(\mu)$ . ■

We show the relationship between Dirichlet forms and Quadratic forms associated with Schrodinger Operators. Let the potential  $V$  satisfy hypothesis (H1). We assume that  $\inf \text{spec}(H) = E$  is an eigenvalue and let  $\psi$  be the corresponding ground state eigenfunction.

Let  $h = -\log \psi$  and define the Borel probability measure  $\mu$  on  $\mathbb{R}$  by

$$d\mu(x) = e^{-2h(x)} dx$$

Define the operator  $\bar{D}$  by

$$(2.14) \quad \bar{D} = C(H-E)C^{-1}$$

where  $C$  is the unitary operator from  $L^2(\mathbb{R}, dx) \rightarrow L^2(\mu)$  defined by

$$(2.15) \quad C\phi = \psi^{-1}\phi, \quad \phi \in L^2(\mathbb{R}, dx).$$

$\bar{D}$  is a positive self-adjoint operator in  $L^2(\mu)$ , 0 is a simple eigenvalue and the constant function 1 is the corresponding eigenfunction.

In fact  $\bar{D}$  is the unique positive self-adjoint operator associated to the closed positive bilinear form  $\bar{\delta}$  corresponding to  $\epsilon-E$  (in the unitary equivalence  $C$ ) and  $C_c^\infty(\mathbb{R})$  is a core for  $\bar{\delta}$ .

Since  $\psi$  is bounded and locally bounded away from 0,  $h$  is bounded and locally bounded above. Since  $\psi \in Q(\epsilon_0)$  and the first order partial derivatives of  $h$  are in  $L_{loc}^2$ , we can associate with  $\mu$  a Dirichlet form  $\delta$  and a Dirichlet Operator  $D$ . Now  $\delta$  is defined as the closure of  $\delta^a$  form whose domain is  $C_c^\infty(\mathbb{R})$  and  $C_c^\infty(\mathbb{R})$  is a form core for  $\bar{\delta}$ , one can show that  $\delta = \bar{\delta}$  and  $D = \bar{D}$  and hence the Dirichlet form and Dirichlet operator are unitarily equivalent to the Schrodinger form and Schrodinger operator.

#### 2.4. Stochastic Processes Associated with Dirichlet Forms and Dirichlet Operators.

The unitary equivalence between Schrodinger and Dirichlet operators exhibited in the previous section has an important role in Nelson's stochastic mechanics and also in non-linear filtering theory.

We consider the Schrodinger operator  $H = -\frac{1}{2}\Delta + V$  defined as a sum of quadratic forms on  $L^2(\mathbb{R}, dx)$ . We assume that  $\inf \text{ spectrum } (H) = E$  (assumed to be 0) is an eigenvalue and the corresponding eigenfunction  $\psi \geq 0$  and normalized  $\int_{\mathbb{R}} \psi(x)^2 dx = 1$ . Define the Borel probability measure  $\mu$  by  $d\mu(x) = (\psi(x))^2 dx$ . Let  $\delta$  be the corresponding Dirichlet form and  $D$  the corresponding Dirichlet operator. Let  $h = -\log \psi$ .

We want to construct a Markov diffusion process which corresponds to a stochastic differential equation with drift  $-\nabla h$ . There is an obvious difficulty in interpreting the stochastic differential equation in a strong sense. But we can construct a weak solution using the measure transformation technique of Girsanov and the Feynman-Kac formula.

To do this we assume the ground state  $\psi(x) > 0, \forall x$  (by choosing a representative from an equivalence class) and we also assume that  $\psi(x) \rightarrow 0$  as  $|x| \rightarrow \infty$ .

Now we use the unitary equivalence between the Dirichlet operator  $D$  and the

operator  $H = -\frac{1}{2} \frac{d^2}{dx^2} + V(x)$  to conclude that

$$(2.16) \quad (e^{-tD}f)(x) = \psi(x)^{-1} E_{W_x} \left[ \exp\left(-\int_0^t V(W_s) ds\right) \psi(W_t) f(W_t) \right]$$

where  $e^{-tD}$  is the semi-group generated by  $-D$ .

For each  $t > 0$ , consider

$$(2.17) \quad L_t = \psi(x)^{-1} \psi(W_t) \exp\left(-\int_0^t V(W_s) ds\right), \text{ which}$$

is a random variable, positive  $W_x$  -a.s. and  $E_{W_x}(L_t) = 1$ . Hence,  $(\Omega, F_t, L_t)$  is a martingale, where  $F_t$  denotes the smallest  $\sigma$ -algebra for which the coordinate functions  $W_t$  are measurable. Hence for each  $x$ ,  $P_x = L_t \cdot W_x \Big|_{F_t}$  is a probability measure on  $(\Omega, F)$ .

Now by explicit calculation,  $V(x) = \frac{1}{2}(-\Delta h(x) + |\nabla h(x)|^2)$  and since  $h = -\log \psi$ , we get from (2.17)

$$(2.18) \quad L_t = \exp[-h(W_t) + h(x) - \frac{1}{2}(\int_0^t \Delta h(W_s) ds - \int_0^t |\nabla h(W_s)|^2 ds)]$$

The idea now is to apply the Ito-differential to the function  $h$ . There is a difficulty here because the function  $h$  does not have continuous partial derivatives upto order two. But  $h$  is a convex function and for these functions the Ito-differential rule can be extended to continuous functions with first partial derivatives in  $L^2_{loc}(\mathbb{R}, dx)$  and second partial derivatives in  $L^1_{loc}(\mathbb{R}, dx)$ . Applying the generalized Ito-differential rule to (2.18), we get

$$L_t = \exp[-\int_0^t \nabla h(W_s) dW_s - \frac{1}{2} \int_0^t |\nabla h(W_s)|^2 ds].$$

Therefore by the Girsanov theorem, the process  $B_t = W_t - W_0 + \int_0^t \nabla h(W_s) ds$  is a  $(F_t, P_x)$  standard Brownian and calling  $W_t = X_t$ , the stochastic process  $(X_t | t \geq 0)$  considered on the probability space  $(\Omega, F_t, P_x)$  is a unique weak solution of

$$(2.19) \quad \begin{cases} dX_t = -\nabla h(X_t) ds + dB_t \\ X_0 = x, \text{ a.s.} \end{cases}$$

Furthermore the measure  $\mu$  defined by  $d\mu(x) = \psi(x)^2 dx$  is the unique finite invariant measure of (2.19).

From the construction of the probability measure  $P_x$ , we see that it has a transition density

$$(2.20) \quad \rho(t, y; 0, x) = \psi(y) \psi^{-1}(x) E_{W_x} \left[ \exp \left( - \int_0^t V(W_s) ds \right) \middle| W_t = y \right] p(t, y; 0, x)$$

where  $p$  is the transition density of Brownian motion.

We remark that it is this transition density we wish to compute since this corresponds to the fundamental solution of a parabolic partial differential equation. We shall see later the importance of (2.20) for non-linear filtering problems. For non-linear filtering problems the decay properties of this transition density are also of importance.

We investigate these matters now.

Firstly, we can check that for  $\forall t > 0$ ,  $\rho(t, y; 0, x)$  is a continuous function of the pair  $(x, y)$ . In fact the following estimate holds:

$$(2.21) \quad \left\{ \begin{array}{l} \exists \text{ constants } c_1, c_2 \text{ such that} \\ \forall t > 0, \forall (x, y) \in \mathbb{R} \times \mathbb{R}, \rho(t, y; 0, x) \leq c_1 e^{c_2 t} \psi(y) \psi^{-1}(x) p(t, y; 0, x) \end{array} \right.$$

If the semigroup

$(T_t f)(x) = E_{W_x} [f(W_t) \exp(-\int_0^t V(W_s) ds)]$  is compact (see Theorem 2.1) and its spectrum is strictly bounded away from zero, then the density  $\rho(t, y; 0, x)$  satisfies:

$$(2.22) \quad \left\{ \begin{array}{l} \exists \text{ positive constants } c_1, c_2 \text{ such that} \\ \sup_{y \in \mathbb{R}} \left| \rho(t, y; 0, x) - \psi(y)^2 \right| \leq c_1 \psi^{-1}(x) e^{-c_2 t}, \forall t > 0, \forall x \in \mathbb{R}. \end{array} \right.$$

## 2.5 Hypercontractive Semigroups

In the previous sections we have seen that for a large class of potentials, the semigroup  $T_t^V$  has a negative infinitesimal generator which coincides with a self-adjoint extension of  $-\frac{1}{2} \frac{d^2}{dx^2} + V(x)$ . In this section we point out that the

semigroups defined by the Feynman-Kac formula are often Hypercontractive semigroups. These semigroups have played an important role in constructive quantum field theory and are likely to play an equally important role in the theory of non-linear filtering.

We follow the notation, hypotheses (specially on the potential  $V$ ) and the terminology of the previous sections.

We consider the Dirichlet semi-group  $(e^{-tD}, t \geq 0)$  on  $L^2(\mu)$  where  $d\mu(x) = \psi(x)^2 dx$ .

Definition: The semigroup  $(e^{-tD}, t \geq 0)$  is said to be hypercontractive if for some  $t > 0$  and some  $r > 2$ ,  $e^{-tD}$  is a bounded operator from  $L^2(\mu)$  into  $L^r(\mu)$ .  $\square$

From our point of view the best approach towards the question of hypercontractivity is via the approach of Gross using Logarithmic Sobolev inequalities.

Definition: The operator  $D$  is called a Sobolev generator if for some real constants  $c > 0$  and  $\gamma$  we have

$$(2.23) \quad \int_{\mathbb{R}} |f|^2 \text{Log} |f| d\mu \leq c(Df, f)_{L^2(\mu)} + \gamma \|f\|_{L^2(\mu)}^2 + \|f\|_{L^2(\mu)}^2 \text{Log} \|f\|_{L^2(\mu)}$$

for all  $f \in \mathcal{D}(D)$ . The constants  $c$  and  $\gamma$  are called the Sobolev coefficient and the local norm of  $D$ .  $\square$

Logarithmic Sobolev Inequality (Gross)

If  $\mu$  is a probability measure on  $\mathbb{R}$  which satisfies:

$$(2.24) \quad \int_{\mathbb{R}} |f|^2 \text{Log} |f| d\mu \leq c \int_{\mathbb{R}} |\nabla f|^2 d\mu + \gamma \|f\|_{L^2(\mu)}^2 + \|f\|_{L^2(\mu)}^2 \text{Log} \|f\|_{L^2(\mu)}$$

for some constants  $c > 0$ ,  $\gamma \geq 0$  and  $\forall f$  which are bounded functions with distributional first order derivative in  $L^2(\mu)$ , then for  $\forall r \in [2, +\infty[$

$$(2.25) \quad \int_{\mathbb{R}} |f|^2 \text{Log} |f| d\mu \leq c(r) \int_{\mathbb{R}} \nabla f \cdot \nabla f_r d\mu + \gamma \|f\|_{L^2(\mu)}^r + \|f\|_{L^2(\mu)}^r \text{Log} \|f\|_{L^2(\mu)}$$

as above, where  $f_r = (\text{sgn} f)|f|^{r-1}$  and  $c(r) = \frac{cr}{2(r-1)}$ .  $\square$

The salient facts about hypercontractive semigroups and Sobolev generators are the following (Carmona)

Proposition 2.4.  $(e^{-tD}|t \geq 0)$  is a hypercontractive semigroup if and only if  $D$  is a Sobolev generator.  $\square$

Proposition 2.5.  $D$  is a Sobolev generator with Sobolev coefficient  $c$  if and only if

$-\text{Log } \psi \leq cD + b$ , the above interpreted as quadratic forms on  $L^2(\mu)$ , for some constant  $b$ .  $\square$

Consider potentials satisfying (H1) which have the further property:

$$\forall x \in \mathbb{R}, \quad a_1 P(x) + b_1 \leq V_1(x) \leq a_2 P(x) + b_2$$

where  $a_1, a_2 > 0$ ,  $b_1, b_2 \in \mathbb{R}$  and  $P$  an even polynomial, and  $\liminf_{|x| \rightarrow \infty} |x|^{-2} P(x) > 0$ .

Proposition 2.6. Schrödinger operators with potentials of the above class generate Dirichlet semigroups which are hypercontractive.  $\square$

Example:

(i) The operator  $D = -\frac{1}{2} \frac{d^2}{dx^2} + \frac{d}{dx}$  generates a hypercontractive semi-group  $e^{-tD}$  on  $L^2(\mathbb{R}, \frac{1}{\sqrt{\pi}} e^{-x^2} dx)$ .

(ii) Consider a stochastic differential equation:

$$dx_t = f(x_t)dt + dw_t$$

Suppose that the Riccati equation

$$\frac{df}{dx} + f^2 = V(x), \quad \text{where } V(x) \text{ is an even polynomial satisfying the hypotheses}$$

of this section such that a global solution exists for the Riccati equation. Then the generator of the associated diffusion process is in fact a Dirichlet operator and generates a hypercontractive semigroup on an appropriate  $L^2(\mu)$ -space.

Notes and References for Section 2

(i) For general references for this section, consult

1. B. Simon: Functional Integration and Quantum Physics, Academic Press, New York, 1979.
2. M. Reed and B. Simon: Methods of Modern Mathematical Physics, Vols. I and II, Academic Press, New York, 1972, 1975.
3. D.W. Stroock and S.R.S. Varadhan: Multi-dimensional Diffusion Processes, Springer-Verlag, Berlin, New York, 1979.

(ii) In this section we follow very closely:

1. R. Carmona: Regularity Properties of Schrödinger and Dirichlet Semigroups, J. of Functional Analysis 33, 259-296 (1979).

(iii) The material in Section 2.4 is apparently partially new and uses the Girsanov transformation to obtain a weak solution of a stochastic differential equation.

For a slightly different approach see:

1. H. Ezawa, J.R. Klauder and L. Shepp: A Path Space Picture for Feynman-Kac Averages, Annals of Physics 88, 588-620 (1974).

(iv) For the generalized Ito Differential rule, see A. Bensoussan and J.L. Lions: Applications des Inéquations Variationnelles en Contrôle Stochastique, Dunod, Paris 1978.

(v) The reference to stochastic mechanics is: Dynamical Theories of Brownian Motion, Princeton University Press, Princeton, N.J., 1967.



### 3. The Bayes Formula and the Unnormalized Conditional Density Equation

Let  $(\Omega, \mathcal{F}, P)$  be a complete probability space,  $\mathcal{F} = (\mathcal{F}_t)$ ,  $0 \leq t \leq 1$  a non-decreasing family of sub  $\sigma$ -algebras of  $\mathcal{F}$  and  $(\eta_t, \mathcal{F}_t)$ ,  $0 \leq t \leq 1$  a Wiener Process. Consider the signal process  $z = (z_t, \mathcal{F}_t)$  and the observation equation

$$(3.1) \quad y_t = \int_0^t z_s ds + \eta_t.$$

Let

$$(3.2) \quad z_t = h(x_t), \text{ and}$$

$x_t$  be the solution of the Ito stochastic differential equation

$$(3.3) \quad dx_t = f(x_t)dt + g(x_t)d\omega_t.$$

We make the following assumptions:

A1. The functions  $f$  and  $g$  are continuous and bounded and equation (3.3) has a unique solution in the weak sense for each initial condition  $x_0$  (later we shall have to make further smoothness assumptions).

A2.  $x_t$  and  $\eta_t$  are independent and  $E(\int_0^1 h(x_s)^2 ds) < \infty$ .

Under the above hypotheses we can write down a Bayes formula for computing  $E(\phi(x_t) | \mathcal{F}_t^Y)$  where  $\mathcal{F}_t^Y$  is the  $\sigma$ -field generated by  $\{y_s | 0 \leq s \leq t\}$  and  $E(\int_0^1 h(x_s)^2 ds) < \infty$ . This relies on the Girsanov transformation. It is known that there is a new equivalent measure  $\tilde{P}$  under which  $y_t$  and  $x_t$  are independent, the distribution of  $x_t$  remains invariant and under this new measure  $y_t$  is Brownian motion. The Radon-Nikodym Derivative of  $P$  with respect to  $\tilde{P}$  is given by

$$\left. \frac{dP}{d\tilde{P}} \right|_{\mathcal{F}_t^n} = \exp\left(\int_0^t h(x_s) dy_s - \frac{1}{2} \int_0^t h^2(x_s) ds\right) \triangleq R_t$$

Hence

$$E(\phi(x_t) | F_t^y) = \frac{\tilde{E}(\phi(x_t) L_t | F_t^y)}{\tilde{E}(L_t | F_t^y)}$$

Using the independence, it is convenient to view this conditional expectation as expectation on a product space  $(\Omega \times \hat{\Omega}, \mathcal{A} \times \hat{\mathcal{A}}, P \otimes P)$  and write it explicitly as

$$(3.4) \quad \hat{\phi}(x_t) = \frac{\int_{\hat{\Omega}} \phi(x_t(\hat{\omega})) \exp(\int_0^t h(x_s(\hat{\omega})) dy_s(\omega) - \frac{1}{2} \int_0^t h(x_s(\hat{\omega}))^2 ds) dP(\hat{\omega})}{\int_{\hat{\Omega}} \exp(\int_0^t h(x_s(\hat{\omega})) dy_s(\omega) - \frac{1}{2} \int_0^t h(x_s(\hat{\omega}))^2 ds) dP(\hat{\omega})}$$

This is the Bayes Formula.

If we replace  $y_s$  by  $x_s$  for  $x \in C(0,1;\mu_x)$  and define

$$(3.5) \quad \rho(t,x,\hat{\omega}) = \exp(\int_0^t h(x_s(\hat{\omega})) dx_s(\omega) - \frac{1}{2} \int_0^t h(x_s(\hat{\omega}))^2 ds) \text{ and}$$

$$(3.6) \quad g(t,x) = \int_{\hat{\Omega}} \rho(t,x,\hat{\omega}) dP(\hat{\omega}) ,$$

then  $g(t,x)$  is the Radon-Nikodym derivative  $\frac{d\mu_y}{d\mu_\eta}(t,x)$ , where  $\mu_y$  is the measure induced on  $C(0,1)$  by the observation and  $\mu_\eta$  is the Wiener measure on  $C(0,1)$  induced by  $\eta$ .

Let  $\sigma_t(\phi)$  denote the numerator of (3.4). It will be also convenient to write the numerator of (3.4) as a Feynman-Kac formula

$$(3.7) \quad \sigma_t(\phi) = \int_X \phi(\xi_t) \exp(\int_0^t h(\xi_s) dy_s - \frac{1}{2} \int_0^t h^2(\xi_s) d\mu_x(\xi)) .$$

and  $X = C(0,1)$ .

Let  $P_t$  be the semigroup of the diffusion process corresponding to (3.3) and  $L_0 = \frac{1}{2} g^2(x) \frac{\partial^2}{\partial x^2} + f \frac{\partial}{\partial x}$  be its infinitesimal generator. Then  $\sigma_t$  satisfies the stochastic partial differential equation

$$(3.8) \quad d\sigma_t(\phi) = \sigma_t(L_0\phi)dt + \sigma_t(L_1\phi)dy_t, \quad \forall \phi \in \mathcal{D}(L_0) \cap \mathcal{D}(L_1), \text{ where}$$

$L_1$  is the unbounded multiplication operator  $h$ .

This is one of the fundamental results of non-linear filtering. Now using the fact that  $h(x_s)$  is a semi-martingale and using the relationship between Ito and Fisk-Stratanovich integrals we get

$\int_0^t \sigma_s(h\phi)dy_s = \int_0^t \sigma_s(h\phi) \cdot dy_s - \frac{1}{2} \langle \sigma(h\phi), y \rangle_t$ , where  $\langle \cdot, \cdot \rangle_t$  denotes the quadratic variation and  $\cdot$  denotes the Fisk-Stratonovich integral and hence we obtain

$$(3.9) \quad d\sigma_t(\phi) = \sigma_t(L_0\phi)dt - \frac{1}{2}\sigma_t(h^2\phi) + \sigma_t(h\phi) \cdot dy_t, \quad \text{which}$$

we write symbolically as

$$(3.10) \quad \frac{d\sigma_t(\phi)}{dt} = \sigma_t[(L_0 + \dot{y}_t h - \frac{1}{2} h^2)(\phi)] \quad , \quad \text{which}$$

we can integrate by the Feynman-Kac formula

$$(3.11) \quad \sigma_t(\phi) = E_{\sigma_0} [\exp(\int_0^t \dot{y}_s h ds - \frac{1}{2} \int_0^t h^2 ds) \phi(x_t)]$$

where  $\sigma_0$  is a measure-valued random variable independent of  $y$  such that  $E[\sigma_0(\phi)] = \mu(\phi)$  and  $\mu$  is the initial distribution.

Define

$$\Sigma(t, x_t, y_0^t) = \tilde{E}[L_t | \sigma\{x_t\}_{xF_t^y}]$$

and define

$$q(t, z, y_0^t) = \Sigma(t, z, y_0^t) p(t, z)$$

where  $p(t, z)$  is the density of the  $x$ -process. Then

$$\sigma_t(\phi) = \int_R \phi(z) q(t, z, y_0^t) dz.$$

Now  $\sigma_t(\phi) = \int \phi d\sigma_t$ , and hence we see that  $\sigma_t$  has a density  $q(t, z, y_0^t)$ , and hence from (3.7)

$$(3.12) \quad q(t, z, y_0^t) = \int_X \exp \left( \int_0^t h(x_s) dy_s - \frac{1}{2} \int_0^t h^2(x_s) ds \right) d\mu^z(x)$$

where the integration is over the path space of  $x$  with  $x_0 = 0$  and  $x_t = z$ . It can be shown that  $q_t \triangleq q(t, z, y_0^t)$  satisfies the stochastic partial differential equation

$$(3.13) \quad dq_t = (L_0^* - \frac{1}{2} L_1^2) q_t dt + L_1 q_t \cdot dy_t,$$

where  $L_0^*$  is the formal adjoint of  $L_0$ .

Equation (3.13) is the Mortensen-Zakai equation for the unnormalized conditional density.

### 3.1 Mortensen-Zakai Equations and Gauge Transformations

The study of non-linear filtering is the study of the Mortensen-Zakai equation--its explicit solution and its group invariance properties. Equivalently it is the study of the function space representation of  $\sigma_t(\phi)$  given in (3.7) and (3.12). In (3.7) we have the stochastic integral  $\int_0^t h(\xi_s) dy_s$ . Under our hypotheses we can write

$$\int_0^t h(\xi_s) dy_s = y_t h(\xi_t) - y_0(\xi_0) - \int_0^t y_s dh(\xi_s).$$

Using the above it is clear that  $\sigma_t$  can be evaluated for all  $y \in C(0,1)$  (not just a subset of full Wiener measure). It has been shown by Clark  $\frac{\sigma_t(\phi, y)}{\sigma_t(1, y)}$  is a version of the conditional expectation  $\hat{\phi}$ .

An equivalent way of considering this problem is to eliminate the stochastic integral from the Mortensen-Zakai equation (3.13).

This can be done by making the observation that the operator  $L_1$  = multiplication by  $h$  is a diagonal operator, and its effect can be removed via a time-dependent Gauge transformation. To see this write  $q_t = \exp(h(x)y_t)\tilde{q}_t$ . Then a direct calculation shows that  $\tilde{q}_t$  satisfies an ordinary partial differential equation (parametrized by  $y$ ):

$$(3.14) \quad \frac{d\tilde{q}_t}{dt} = \exp(-h(x)y_t)(L_0^* - \frac{1}{2}L_1^2)(\exp(h(x)y_t)\tilde{q}_t)$$

It is interesting to rewrite equation (3.14) in a form which brings out the commutation properties of  $L_0^* - \frac{1}{2}L_1^2$  and  $L_1$ . If we denote by  $L_2 = [L_0^* - \frac{1}{2}L_1^2, L_1]$  and  $L_3 = [L_1, L_2]$  considered as formal differential operators and computed on some common invariant domain, (3.13) can be rewritten as:

$$(3.15) \quad \frac{d\tilde{q}_t}{dt} = (L_0^* - \frac{1}{2}L_1^2)\tilde{q}_t + y_t L_2 \tilde{q}_t - y_t^2 L_3 \tilde{q}_t$$

Explicitly  $L_2 = h_x \frac{d}{dx} + (\frac{1}{2} h_{xx} - f h_x)$ ,  $L_3 = -h_x^2$  (assuming  $g \equiv 1$  for simplicity).

When we are dealing with unbounded observation operators  $h$  it is this equation which is the easiest to deal with. The above also shows that the commutators  $L_2$  and  $L_3$  have an important role to play in the understanding of equation (3.14).

In (3.12) we see that evaluating  $q_t$  involves an integration over the path space of the  $x$ -process. The integration would be simplified if this could be done with respect to Wiener measure. This corresponds to removing the drift term in the operator  $L_0^*$ . For simplicity assume  $g \equiv 1$ . Define the operator multiplication by  $\psi(x)$ , where  $\psi$  is invertible from the intertwining relationship.

$(\frac{d}{dx} - f)\psi = \psi \frac{d}{dx}$ , which defines  $\psi(x) = e^{\int_0^x f(z)dz}$ . Then the operator  $L_0^*$  transforms to  $\psi^{-1} L_0^* \psi = \frac{1}{2} \frac{d^2}{dx^2} - \frac{1}{2} V(x)$ , where  $V(x) = (f_x + f^2)$ . Then  $\hat{q}_t = \psi^{-1} q_t$  is the solution of

$$(3.16) \quad d\hat{q}_t = \hat{L}_0 \hat{q}_t dt - \frac{1}{2} V(x) \hat{q}_t dt + h \hat{q}_t \cdot dy_t, \text{ where } \hat{L}_0 = \frac{1}{2} \frac{d^2}{dx^2} - \frac{1}{2} h^2.$$

This is also an example of a Gauge transformation. Equation (3.15) involves the (imaginary time) Schrödinger operator  $-\frac{1}{2} \frac{d^2}{dx^2} + \frac{1}{2} h^2 + \frac{1}{2} V(x)$ . It should be noted that the Lie Algebra of operators  $LA\{L_0^* - \frac{1}{2} L_1^2, L_1\}$  and  $LA\{\hat{L}_0 - V(x), L_1\}$  are isomorphic. Recall in section 2 we have seen how certain Schrödinger operators are unitarily equivalent to Dirichlet operators. (See later section also.)

### 3.2 Integration of the Mortensen-Zakai Equation and Calculus of Variations

In section 3 we have seen that solving the Mortensen-Zakai equation is equivalent to evaluating the function-space integral

$$q(t, z, y_0^t) = \int_X \exp(\int_0^t h(x_s) dy_s - \frac{1}{2} \int_0^t h^2(x_s) ds) d\mu^z(x).$$

We now show how this can be done by adapting certain ideas of Feynman. Feynman's idea was to separate the classical and quantum parts of the total quantum motion and also to understand the relationship between classical and quantum mechanics. We shall attempt to separate the contribution of the observation in  $q_t$  from that of the signal process. For the class of problems considered by Benes this can be done explicitly.

For this purpose, we consider that the  $x_t$ -process is governed by

$$dx_t = f(x_t)dt + dw_t, \text{ and the observation equation is}$$

$$dy_t = x_t dt + d\eta_t$$

$$\text{Hence } L_0^* = \frac{1}{2} \frac{d^2}{dx^2} - \frac{d}{dx} f, \text{ and } L_1 = \text{mult. by } x.$$

We do a gauge transformation and attempt to solve

$$d\hat{q}_t = \hat{L}_0 \hat{q}_t dt - V(x) \hat{q}_t dt + x \hat{q}_t dt, \text{ where } V(x) = \frac{1}{2}(f_x + f^2)$$

Hence we have to compute:

$$(3.17) \quad \hat{q}(t, z, y_0^t) = E_{W^z} [\exp(\int_0^t x_s dy_s - \frac{1}{2} \int_0^t x_s^2 ds - \int_0^t V(x_s) ds)]$$

where the integration over Wiener-measure with  $x_0 = 0$  and  $x_t = z$ .

We show that if  $V(x) = ax^2 + bx + c$ , and  $f_x + f^2 = V(x)$  has a global solution then a sufficient statistic for computing  $\hat{q}$  is the set of equations:

$$(3.18) \quad \begin{cases} d\xi_s = -\eta_s ds & ; \quad \xi_0 = 0, \xi_t = z \\ d\eta_s = -\xi_s ds - V_\xi(\xi_s) ds + dy_s \end{cases} \quad (1)$$

We call the above set of equations the bi-characteristics of the Mortensen-Zakai equation. To see this we introduce the transformation

$$dx_s = d\xi_s + dz_s, \text{ where } \xi_s \text{ is given by the above and we do not impose the boundary conditions.}$$

---

(1) This stochastic two-point boundary value problem has to be given an appropriate sample-pathwise interpretation.

By essentially the Girsanov Theorem,

$$\frac{d\mu_z}{d\mu_x} = \exp\left(\int_0^t \eta_s dz_s - \frac{1}{2} \int_0^t \eta_s^2 ds\right) \quad (1)$$

Now impose the boundary conditions and hence in terms of the z-variable

$$\hat{q}(t, z, y_0^t) = \int \exp\left(\int_0^t (z_s + \xi_s) dy_s - \frac{1}{2} \int_0^t (z_s + \xi_s)^2 ds - \int_0^t V(z_s + \xi_s) ds\right) \\ \times \exp\left(\int_0^t \eta_s dz_s - \frac{1}{2} \int_0^t \eta_s^2 ds\right) d\mu_z^0, \quad ,$$

where the integration is over the path space of z with  $z_0 = 0$  and  $z_t = 0$ .

By applying Ito's rule several times, and using the fact that  $V(x) = ax^2 + 2bx + c$ , we obtain

$$\hat{q}(t, z, y_0^t) = \exp\left(\frac{1}{2} \int_0^t \xi_s dy_s - \int_0^t b \xi_s ds + \frac{1}{2} (z \eta_t)\right) \\ (3.19) \quad \times \int \exp\left(-\int_0^t (V(z_s) + \frac{1}{2} z_s^2) d\mu_z^0\right) \quad .$$

The terms involving the z-process can be explicitly evaluated.

It is worth remarking that (3.18) is the solution of a variational problem.

To see this let us first interpret equation (3.18) in the Stratanovich sense.

Then equation (3.18) are the necessary conditions of the following optimal control problem:

$$\text{Min}_u \frac{1}{2} \int_0^t u_s^2 ds + \frac{1}{2} \int_0^t [(\xi_s - \dot{y}_s)^2 + 2V(\xi_s)] ds \\ \dot{\xi}_s = u_s \quad ; \quad \xi_0 = 0 \quad , \quad \xi_t = z \quad . \quad (2)$$

---

(1) By conditioning.

(2) This variational problem can be given rigorous meaning using recent work of Bismut.



If  $y_t \equiv 0$  and  $V \equiv 0$ , then the equations (3.18) correspond to the imaginary time harmonic oscillator and our methods show the relation between a euclidean quantum harmonic oscillator and an imaginary time classical harmonic oscillator.

These methods could also be applied to study the limiting behavior of the Cauchy problem

$$\frac{\partial \psi}{\partial t}(t, x) = \lambda \nabla_x^2 \psi + \frac{V(x)}{\lambda} \psi \quad \text{as } \lambda \rightarrow 0$$

with an initial condition  $\psi_\lambda(x, 0) = \exp[-S_0(x)/\lambda]$  and show that the limiting solution satisfies the corresponding equations of classical mechanics. This would provide an alternative derivation to the results of Maslov who treated the Schrödinger equation version of this problem.

Equations (3.18) is related to the smoothing problem. It turns out, that the unnormalized conditional density could also be evaluated by the following sufficient statistic.

$$(3.20) \quad \begin{cases} d\alpha_s = (f(\alpha_s) - \beta_s)ds & ; \quad \alpha_0 = 0 \quad \alpha_t = z \\ d\eta_s = -(f_\alpha(\alpha_s)\eta_s + \alpha_s)ds + dy_s \end{cases} .$$

These are the Euler-Lagrange equations for the following optimal control problem:

$$\text{Min}_u \frac{1}{2} \int_0^t u_s^2 ds + \frac{1}{2} \int_0^t \{(\alpha_s - \dot{y}_s)^2 + 2V(\alpha_s)\} ds$$

$$\dot{\alpha}_s = f(\alpha_s) + u_s \quad ; \quad \alpha_0 = 0 \quad \alpha_t = z .$$

For the general Kalman filtering problem

$$(3.21) \quad \begin{cases} dx_t = Fx_t dt + Gdw_t \\ dy_t = Hx_t dt + d\eta_t \end{cases},$$

the equations corresponding to (3.20) would be

$$(3.22) \quad \begin{cases} d\alpha_t = F\alpha_t dt - GG'\beta_t dt ; & \alpha_0 = 0 \quad \alpha_t = z \\ d\beta_t = -H'H\alpha_t dt - F'\beta_t dt + H'dy_t \end{cases}$$

These can be recognized to be the smoothing equations given in Hamiltonian form. Our methods show that these equations are intrinsically attached to the Mortensen-Zakai equation and play the role of bi-characteristics corresponding to Hamiltonian-Jacobi equations.

Although we do not do it in this paper it seems reasonable to believe that a perturbation theory analogous to Maslov's work could be carried out for non-linear filtering using the framework used in this section (see forthcoming work of Duncan and Mitter).

Notes and References for Section 3

- (i) For the derivation of the Bayes formula and the Mortensen-Zakai equations see E. Wong: Stochastic Processes in Information and Dynamical Systems, McGraw Hill, New York, 1971 and the references cited there.
- (ii) The seminal paper on non-linear filtering is: M. Fujisaki, G. Kallianpur and H. Kunita: Stochastic Differential equations for the Non-linear Filtering Problem, Osaka J. of Math., Vol. 9, 1972, pp 19-40.
- (iii) Equation (3.14) was first derived by J.M.C. Clark: The Design of Robust Approximations to the Stochastic Differential Equations of Non-linear Filtering in Communication Systems and Random Process Theory: ed. J.K. Skwirzynski, Sithoff and Noordhoff, 1978. For more recent work see for example M.H.A. Davis: A Path-wise Solution of the Equations of Non-linear Filtering, to appear. Writing it in a form involving the commutators makes it clear that for certain problems these equations can be integrated using group invariance methods.
- (iv) The idea of using gauge transformations in the context of non-linear filtering theory is new, although it is implicit in the work of Benes. See also R.W. Brockett: On the Invariance Group of the Conditional Density Equations, Proceedings of the IEEE Decision and Control Conference, 1979, Ft. Lauderdale, Florida. The gauge transformation can be introduced in a much more general setting.
- (v) The variational interpretation of certain non-linear filtering problems in the form presented in Section 3.2 is new and uses certain ideas of Feynman. See for example: R.P. Feynman and A.R. Hibbs: Quantum Mechanics and Path Integrals, McGraw Hill, New York, 1965. For the Girsanov transformation and absolute continuity of measures, see R.S. Lipster and A.N. Shiriyayer: Statistics of Random Processes I, Springer-Verlag, New York, 1977.
- (vi) The reference to Maslov's work is: V.P. Maslov: Théorie des Perturbations et Méthodes Asymptotiques, Dunod-Gauthier-Villars, Paris, 1972.

#### 4. Multiple Ito Integrals and Fock Space

There is a close relationship between the theory of multiple Ito integrals and Wick polynomials. These objects also have an important role to play in Wiener's theory of homogeneous chaos and representations of the Weyl Commutation relations on Fock space. These ideas and constructions also are of importance in non-linear filtering theory.

##### 4.1 Multiple Ito Integrals

The multiple Ito integral of order  $K$  is a map  $f \mapsto I_K(f): L^2(\mathbb{R}^K) \rightarrow L^2(\Omega, \mathcal{F}, P)$  having the following properties:

$$(i) \quad I_K(h) = \prod_{i=1}^K W(A_i) \text{ if } h = 1_{A_1} \times \dots \times 1_{A_K} \text{ for}$$

disjoint rectangles  $A_1, \dots, A_K$ .

$$(ii) \quad I_K(f+g) = I_K(f) + I_K(g).$$

$$(iii) \quad \text{If } f_j \rightarrow f \text{ in } L^2(\mathbb{R}^K) \text{ then } I_K(f_j) \rightarrow I_K(f) \text{ in } L^2(\Omega, \mathcal{F}, P).$$

If  $k = 0$  and  $h \in \mathbb{R}$ , define  $I_0(h) = h$ .

Let  $\tilde{f}(t_1, \dots, t_K) = \frac{1}{K!} \sum_{\pi} f(t_{\pi(1)}, \dots, t_{\pi(K)})$  denote the symmetrization of  $f$ . The mapping  $f \mapsto \tilde{f}$  is the projection of  $L^2(\mathbb{R}^K)$  onto the subspace  $L_s^2(\mathbb{R}^K)$  of  $L^2(\mathbb{R}^K)$  spanned by the symmetric functions.

The multiple Ito integral has the following properties:

$$(a) \quad \text{For } f \in L^2(\mathbb{R}^K), g \in L^2(\mathbb{R}^{K'}) \text{, } I_K(f) = I_K(\tilde{f}) \text{ and}$$

$$E(I_K(f)I_{K'}(g)) = 1_{\{K=K'\}} K! \int_{L^2} (\tilde{f}, \tilde{g})$$

$$(b) \quad \text{For } \phi \in L^2(\mathbb{R}) \text{ and } \lambda \in \mathbb{C},$$

$$\exp(\lambda \int_{\mathbb{R}} \phi_s dw_s - \frac{1}{2} \lambda^2 \int_{\mathbb{R}} \phi_s^2 ds) = \sum_{k=0}^{\infty} \frac{\lambda^k}{k!} \int_{\mathbb{R}^k} \phi_{s_1} \dots \phi_{s_k} dw_{s_1} \dots dw_{s_k}$$

$$(c) \quad L^2(\Omega, F, P) = \bigoplus_{K=0}^{\infty} \{I_K(r) \mid r \in L^2(\mathbb{R}^K)\} \cong \bigoplus_{K=0}^{\infty} L^2_S(\mathbb{R}^K) .$$

We now consider two applications of multiple Ito integrals

#### 4.2 A Representation Theorem for the Best Estimate of a Signal

Consider the non-linear filtering problem

$$y_s = \int_0^t z_s ds + \eta_s ,$$

and assume that the hypotheses of Section 3 (assumptions A1 and A2) hold. Suppose that we want a representation for  $\hat{z}_s = E(z_s | F_t^y)$ . It is a standard result that the innovations process

$$v_t = y_t - \int_0^t \hat{z}_s ds .$$

It is well known that  $v_t$  is standard Brownian motion. Recently we have shown

$F_t^y = F_t^v \text{ mod } P$ . Hence  $\hat{z}_t \in L^2(\Omega, F_t^v, P)$  and hence from property (c)

$$\hat{z}_t = z_0 + \int_0^t k_1(t,s) dv_s + \int_0^t \int_0^{s_1} k_2(t,s_1,s_2) dv_{s_1} dv_{s_2} + \dots$$

and a standard application of the definition of conditional expectation as a projection shows

$$k_1(t,s) = \frac{\partial}{\partial s} E(x_t v_s) , \quad k_2(t,s_1,s_2) = \frac{\partial^2}{\partial s_1 \partial s_2} E(x_t v_{s_1} v_{s_2}) \quad \text{etc.}$$

This representation is not too useful since the computation of the innovations requires computing  $\hat{z}_t$ . However for the following problem it immediately leads to a finite dimensional filter. Suppose that  $z_s = Hx_s$  and  $dx_s = Fx_s ds + Gdw_s$ . We are required to estimate  $x_s$  and  $\alpha_s = P(x_s, 0 \leq s \leq t)$  where  $P$  is a polynomial functional of  $x_s$  with separable kernels. Then it is not too difficult to show that the estimate is also a polynomial functional with separable kernels of the innovations  $dv_s = dy_s - H\hat{x}_s ds$  where  $\hat{x}_s$  is the Kalman filter estimate.

The ideas of multiple Ito integrals also have applications to representing the density  $q(t, z, y_0^t)$  as a multiple integral expansion in the semi-martingale  $y_s$ .

#### 4.2 Random Fields

In quantum field theory random fields (weak distributions) and their polynomial functions (suitably defined) have played an important role. A Wiener integral or more generally an Ito integral is an example of a random field. Multiple Ito integrals can be considered to be polynomial functions of random fields, provided these polynomial functions are suitably renormalized. These are the Wick polynomials. Orthogonal polynomials and in particular the Hermite polynomials have a special role to play in this theory.

#### 4.2.1 Basic Definitions and Preliminaries

Let  $H$  be a real Hilbert space, with the scalar product  $\langle \cdot, \cdot \rangle$ . We shall identify  $H$  with its topological dual. Let  $(\Omega, \mathcal{A}, \mu)$  be a probability space. We denote by  $L^0(\Omega, \mathcal{A}, \mu)$  the space of real (complex)-valued random variables on  $(\Omega, \mathcal{A}, \mu)$ .

Definition 4.1. A continuous <sup>(1)</sup> linear function

$$(4.1) \quad F : H \rightarrow L^0(\Omega, \mathcal{A}, \mu)$$

is termed a random field or weak distribution. Two random fields  $F_1$  and  $F_2$  are equivalent if for any  $\{f_1, \dots, f_n\} \subseteq H$ , the joint distribution of  $F_1(f_1), \dots, F_1(f_n)$  and  $F_2(f_1), \dots, F_2(f_n)$  are the same.  $\square$

An example of a random field which is of importance to us is one which is "generated" by the Wiener process. Let  $W(t, \omega)$  denote the Wiener process for  $t \geq 0$ . Define  $W(-t, \omega) = W(t, \omega)$ ,  $t \geq 0$  and thus extend the definition to all of  $\mathbb{R}$ . Then if we define

$$(4.2) \quad F : L^2(\mathbb{R}) \rightarrow L^0(\Omega, \mathcal{A}, \mu) \text{ by}$$

$F(f) = \int f dW(t, \omega)$ , where the right hand side is a Wiener integral, then  $F$  is a random field.

Definition 4.2. Given a random field  $F : H \rightarrow L^0(\Omega, \mathcal{A}, \mu)$ , the mean functional is the map.

$$(4.3) \quad M_F(f) = E(F(f)) = \int F(f) d\mu.$$

The mean functional need not exist for all  $f \in H$ . If it exists, then  $M_F$  is a linear (not necessarily continuous) functional on  $H$ .

---

(1) Continuous means if  $f_i \rightarrow f$  in  $H$  then  $F(f_i) \rightarrow F(f)$  in probability.

The covariance functional  $C_F : H \times H \rightarrow \mathbb{C}$  is defined by

$$(4.4) \quad C_F(f, g) = E\{[F(f) - M_F(f)][F(g) - M_F(g)]\}$$

where  $-$  denotes complex conjugation. If  $C_F$  exists for all  $f, g \in H$  then  $C_F$  is a positive Hermitian bilinear form on  $H$ . The variance functional  $V_F$  is defined by  $V_F(f) = C_F(f, f)$  and the correlation functional by  $B_F(f, g) = E[F(f)F(g)]$ .

The random field is said to be bounded if there exists a  $k > 0$ , such that

$$\sqrt{B_F(f, f)} \leq k \|f\| \quad \forall f \in H.$$

The characteristic functional  $\phi_F : H \rightarrow \mathbb{C}$  is defined by

$$(4.5) \quad \phi_F(f) = E[e^{iF(f)}] = \int e^{iF(f)} d\mu.$$

It is a continuous, positive definite functional and  $\phi_F(0) = 1$ .  $\square$

Remark: Bochner's theorem extends to this infinite dimensional situation.  $\square$

The following proposition is a consequence of the Riesz representation theorem.

Proposition 4.1. Suppose  $F$  is a bounded random field. Then there exists a unique vector  $f_F \in H$  (the mean) and unique bounded self-adjoint non-negative operators  $R_F$  (covariance) and  $S_F$  (correlation) such that  $M_F(f) = \langle f, f_F \rangle$ ,  $C_F(f, g) = \langle R_F f, g \rangle$  and  $B_F(f, g) = \langle S_F f, g \rangle$ .  $\square$

Definition 4.3. A random field  $N$  such that  $B_N(f, f) < \infty$  with characteristic function

$$(4.6) \quad \phi_F(f) = \exp[iM_N(f) - \frac{1}{2} V_N(f)]$$

is termed a Gaussian random field.



We shall be interested in isonormal random fields  $N_c$  with mean zero variance parameter  $c > 0$ . Its covariance function is  $C_N(f,g) = c\langle f,g \rangle$ . If  $c = 1$  then we get the unit isonormal random field. Its characteristic function

$$\phi_N(f) = e^{-\frac{\|f\|^2}{2}}.$$

Now let us consider a bounded Gaussian random field  $N$ . Its characteristic function can be written as:

$$\phi_N(f) = \exp[i\langle f, f_N \rangle - \frac{1}{2}\langle R_N f, f \rangle].$$

In particular, if the bounded Gaussian random field is the isonormal field with parameter  $\sigma > 0$  then

$$R_N = \sigma I.$$

Let  $\mathcal{B}(H)$  denote the Borel  $\sigma$ -algebra of  $H$  and let  $\mu$  be a probability measure on  $\mathcal{B}(H)$ . Define the random field  $F : H \rightarrow L^0(H, \mathcal{B}(H), \mu)$  by  $[F(f)](g) = \langle f, g \rangle$ .

In this way every Borel Probability measure  $\mu$  on  $\mathcal{B}(H)$  produces a random field.

Suppose  $F : H \rightarrow L^0(\Omega, \mathcal{A}, \nu)$  is a random field. Then we say  $F$  is generated by  $\mu$  on  $\mathcal{B}(H)$  if  $F$  is equivalent to  $F$ . We then have

Proposition 4.2. A Gaussian random field  $N$  is generated by a probability measure on  $(H, \mathcal{B}(H))$  if and only if the covariance operator  $R_N$  is trace class.  $\square$

Thus if  $\dim H = \infty$ , the isonormal random field with covariance operator  $\sigma I$ ,  $\sigma > 0$  could never be produced by a countably-additive probability measure on  $(H, \mathcal{B}(H))$ .

By the Gross construction of "lifting", the finitely additive measure corresponding to the isonormal field can be "extended" to a countably additive measure on a separable Banach space.  $\square$

$S(\mathbb{R}^n)$  will denote the Schwartz space of rapidly decreasing functions, and  $S'(\mathbb{R}^n)$  its dual, the space of tempered distributions. In the sequel it will be often convenient to define a random field as a mapping  $F : V \rightarrow L^0(\Omega, \mathcal{A}, \mu)$  where  $V$  is a topological space. In particular we shall have occasion to take  $V = S(\mathbb{R})$ . We shall often make the assumption that the random field is determined by  $f \in V$  in the sense that the smallest  $\sigma$ -algebra w.o. to which the random variables  $\{F(f) | f \in V\}$  is measurable is  $\mathcal{A}$ . We call such a random field full.

#### Construction of the Unit Gaussian Random Field.

Let  $H$  be a separable Hilbert Space and let  $(e_n)$  be an orthonormal basis for  $H$ . Let  $\bar{\mathbb{R}}$  be the one point compactification of  $\mathbb{R}$  and let  $\Omega = \prod \bar{\mathbb{R}}$  be the Cartesian countable product of copies of  $\bar{\mathbb{R}}$ .  $\Omega$  is a Compact Hausdorff space in the Tychonov topology. Let  $C(\Omega)$  be the set of continuous functions on  $\Omega$  and let  $P(\Omega)$  be the set of functions  $\phi(x_1, \dots, x_n)$  in  $C(\Omega)$  which depend only on a finite number of copies. With the supremum norm topology  $C(\Omega)$  is a Banach space and (by the Stone-Weierstrass theorem  $P(\Omega)$  is dense in  $C(\Omega)$ ). For  $\phi \in P(\Omega)$  define

$$k(\phi) = (2\pi)^{-\frac{n}{2}} \int \phi(x_1, \dots, x_n) e^{-\frac{\|x\|^2}{2}} \mathbb{R}^n dx_1 \dots dx_n$$

Then  $k : P(\Omega) \rightarrow \mathbb{R}$  is linear and  $|k(\phi)| \leq \|\phi\|$  and hence  $k$  can be extended to a continuous linear functional on  $C(\Omega)$ . By the Fiesz Representation theorem there exists a Borel probability measure  $\mu$  on  $\mathcal{B}(C)$  such that  $k(\phi) = \int \phi d\mu, \forall \phi \in C(\Omega)$

Let  $F(e_n) : \Omega \rightarrow \mathbb{R}$  be multiplication by  $x_n$ ,  $n = 1, 2, \dots$ . Then  $F(e_n)$  is a.e. in  $\mathbb{R}$  and hence  $F(e_n) : \Omega \rightarrow \mathbb{R}$ , measurable. It can be shown that  $F(e_n) \in L^2(\Omega, \mathcal{B}(C), \mu)$  and  $\{F(e_n) | n=1, 2, \dots\}$  generate  $\mathcal{B}(C)$ . If  $f \in H$  and  $f = \sum a_n e_n$  then  $\sum a_n F(e_n)$  converges in  $L^2(\Omega, \mathcal{B}(C), \mu)$  to an element  $F(f)$ . This is the unit full Gaussian Random field on  $H$ .

#### 4.2.2 Random Fields on $S(\mathbb{R}^n)$ and White Noise

Let  $F : S(\mathbb{R}^n) \rightarrow L^0(\Omega, \mathcal{A}, P)$  be a random field. The derivative  $F'$  of the random field is defined by

$$F'(f) = -F(f')$$

$F'$  is a random field and if  $F$  is Gaussian,  $F'$  is also Gaussian. Now a random field  $F : S(\mathbb{R}^n) \rightarrow L^0(\Omega, \mathcal{A}, \mu)$  is generated by a stochastic process (assumed to have square integrable sample paths)  $X : \mathbb{R}^n \times \Omega \rightarrow \mathbb{R}$  if

$$[F(f)](\omega) = \int f(x)X(x, \omega)dx, \quad \forall f \in S(\mathbb{R}^n).$$

Let  $W(t, \omega)$ ,  $t \in [0, \infty]$  be the standard Wiener process and let  $F$  be the Gaussian random field generated by  $W(t, \omega)$ . The covariance functional of  $F$  can be computed as

$$\begin{aligned} C_F(f, g) &= \int_0^\infty \int_0^\infty \min(s, t) f(s)g(t)dsdt \\ &= \int_0^\infty [\hat{f}(s) - \hat{f}(\infty)][\hat{g}(t) - \hat{g}(\infty)]dsdt \end{aligned}$$

where  $\hat{f}(t) = \int_0^t f(s)ds$  and  $\hat{g}(t) = \int_0^t g(s)ds$ .

$F'$  the derivative of  $F$  exists (in the sense of distributions) and its covariance functional is given by

$$C_{F'}(f, g) = C_F(f', g') = \int_0^\infty f(t)g(t)dt = \langle f, g \rangle$$

where  $\langle, \rangle$  represents the natural inner product on  $S(\mathbb{R})$  embedded in  $L^2(\mathbb{R})$ .

$F'$  is white noise.

### 4.3 Second Quantization (After Segal and Nelson)

Let  $H$  be a real Hilbert space and let  $F : H \rightarrow L^0(\Omega, \mathcal{A}, \mu)$  be the unit Gaussian determined random field. If  $f_1, \dots, f_n$  are orthonormal in  $H$  and  $\phi$  is a Bounded Baire function on  $\mathbb{R}^n$ , then

$$\int_{\Omega} \phi(F(f_1), \dots, F(f_n)) = \frac{1}{(2\pi)^{n/2}} \int_{\mathbb{R}^n} \phi(x) e^{-\frac{||x||^2}{2}} dx$$

For concreteness  $(\Omega, \mathcal{A}, \mu)$  may be chosen to be countably infinite copies of  $(\mathbb{R}, \mathcal{B}(\mathbb{R}), (2\pi)^{-1/2} e^{-x^2/2} dx)$ .

If  $E$  denotes expectation on  $(\Omega, \mathcal{A}, \mu)$  then

$$(4.3.1) \quad E(F(f_1) \dots F(f_{2n+1})) = 0$$

$$(4.3.2) \quad E(F(f_1) \dots F(f_{2n})) = \sum \langle f_{i_1}, f_{j_1} \rangle \dots \langle f_{i_n}, f_{j_n} \rangle$$

where the sum is over all pairings of  $1, \dots, 2n$  i.e.  $i_1 < \dots < i_n; j_1 < \dots < j_n$ , and  $(i_1, j_1, \dots, i_n, j_n)$  is a permutation  $1, \dots, 2n$ .

$L^2(\Omega, \mathcal{A}, \mu)$  is denoted by  $L^2(H)$  and  $\Gamma(H)$  denotes  $L^2(H)$ .  $\Gamma(H)_{\leq n}$  be the closed linear span in  $\Gamma(H)$  of all elements of the form  $F(f_1) \dots F(f_m)$  with  $m \leq n$  and let  $\Gamma(H)_n$  denote the orthogonal complement of  $\Gamma(H)_{\leq n-1}$  in  $\Gamma(H)_{\leq n}$ . For  $f_1, \dots, f_n$  in  $H$  we define the Wick polynomial

$$: F(f_1) \dots F(f_n) :$$

to be the orthogonal projection of  $F(f_1) \dots F(f_n)$  into  $\Gamma(H)_n$ . In the special case, where  $H$  is one dimensional and hence  $\Gamma(H) = L^2(\mathbb{R}, \mathcal{B}(\mathbb{R}), (2\pi)^{-1/2} e^{-x^2/2} dx)$ ,  $\Gamma(H)_n$  is the one dimensional subspace spanned by the  $n$ th Hermite polynomial and  $:x^n:$  is the  $n$ th Hermite polynomial normalized so that the leading coefficient is 1.

We have the formula

$$(4.3.3) \quad \langle :F(f_1) \dots F(f_n):, :F(g_1) \dots F(g_n): \rangle = \sum_{\pi} \langle f_{\pi(1)}, g_1 \rangle \dots \langle f_{\pi(n)}, g_n \rangle.$$

where the sum is over all permutations  $\pi$  of  $1, \dots, n$ . If all the  $f$ 's and  $g$ 's are equal, we get

$$(4.3.4) \quad \langle :F(f)^n: , F(f)^n \rangle = \frac{1}{\sqrt{2}} \int_{-\infty}^{\infty} (:x^n:)^2 e^{-x^2/2} dx = n!$$

Let  $H_1$  be the complexification of  $H$  and let  $H_n$  denote the  $n$ -fold Hilbert space symmetric tensor product of  $H_1$  with itself. On  $H_n$  we define the inner product such that

$$(4.3.5) \quad \langle \text{Sym}(f_1 \otimes \dots \otimes f_n), \text{Sym}(g_1 \otimes \dots \otimes g_n) \rangle = \sum_{\pi} \langle f_{\pi(1)}, g_1 \rangle \dots \langle f_{\pi(n)}, g_n \rangle$$

where

$$(4.3.6) \quad \text{Sym}(f_1 \otimes \dots \otimes f_n) = \frac{1}{n!} \sum_{\pi} f_{\pi(1)} \otimes \dots \otimes f_{\pi(n)}$$

From (4.3.3) and (4.3.5), we see that the mapping  $:F(f_1) \dots F(f_n): \mapsto \text{Sym}(f_1 \otimes \dots \otimes f_n)$

extends uniquely to a unitary operator from  $\Gamma(H)_n$  onto  $H_n$ . We use this mapping to identify  $\Gamma(H)_n$  and  $H_n$ . Analogous to the fact that the one-dimensional Hermite polynomials span  $L^2(\mathbb{R}, \mathcal{B}(\mathbb{R}), (2\pi)^{-1/2} e^{-x^2/2} dx)$ , Segal proved

$$(4.3.7) \quad \Gamma(H) = \sum_{n=0}^{\infty} H_n, \text{ for arbitrary real Hilbert space } H.$$

$\Gamma(H)$  is Fock Space.

If the random field  $F(f) = \int f dW$ , where  $f \in L^2(\mathbb{R}) = H$  and  $W$  is the standard Wiener process, then elements of  $\Gamma(H)_n$  are multiple Wiener integrals (in the sense of Ito).

The space  $\Gamma(H)$  is intrinsically attached to the structure of  $H$  as a real Hilbert space. Thus if  $U : H \rightarrow K$  is an orthogonal mapping of one real Hilbert space into another, it induces a unitary mapping  $\Gamma(U) : \Gamma(H) \rightarrow \Gamma(K)$ , where on  $H_n$ ,  $\Gamma(U) = \underbrace{U \otimes \dots \otimes U}_{n\text{-fold}}$ . Similarly if  $I : H \rightarrow K$  is an isometric embedding then it induces an isometric embedding

$\Gamma(I) : \Gamma(H) \rightarrow \Gamma(K)$  and similarly for an orthogonal projection  $E : H \rightarrow K$ .

If  $A : H \rightarrow K$  is a contraction then  $\Gamma(A) : \Gamma(H) \rightarrow \Gamma(K)$  is defined to be the direct sum of  $\Gamma(A)_n : H_n \rightarrow K_n$ , where  $\Gamma(A)_n = \underbrace{A \otimes \dots \otimes A}_{n\text{-fold}}$ . Now any contraction  $A : H \rightarrow K$  can be decomposed as

$$\begin{array}{ccccc} H & \xrightarrow{I} & H \otimes K & \xrightarrow{U} & K \otimes H \\ & \searrow A & & \swarrow E & \\ & & K & & \end{array}$$

where  $I$ ,  $U$  and  $E$  are as above.

Hence  $\Gamma(A) = \Gamma(E)\Gamma(U)\Gamma(I)$ . Now  $\Gamma(A)$  is doubly Markovian in the sense that

$$(4.3.8) \quad \left\{ \begin{array}{l} \alpha \geq 0 \implies \Gamma(A)\alpha \geq 0 \\ \Gamma(A)1 = 1 \\ E\Gamma(A)\alpha = E\alpha, \end{array} \right.$$

Any doubly Markovian operator is a contraction from  $L^p$  to  $L^p$ .

It turns out that  $\Gamma(A)$  has stronger contractive properties and the precise statement of this is an important theorem of Nelson.

Theorem 4.3.1 (Nelson Hypercontractivity Theorem).

Let  $A : H \rightarrow K$  be a contraction. Then  $\Gamma(A)$  is a contraction from  $L^q(H) \rightarrow L^p(K)$  for  $1 \leq q \leq p \leq \infty$  provided that

$$(4.3.9) \quad \|A\| \leq \sqrt{\frac{q-1}{p-1}}$$

If (4.3.9) does not hold then  $\Gamma(A)$  is not a bounded operator from  $L^q(H) \rightarrow L^p(K)$ . ■

#### 4.4 Absolute Continuity of Gaussian Random Fields

Let  $\phi$  be a Gaussian unit random field  $H \rightarrow L^2(\Omega, A, \mu) = \Gamma(H)$ . We consider the following related questions:

(1) Given  $A : H \rightarrow H$  linear, continuous when is there a unitary map

$U : \Gamma(H) \rightarrow \Gamma(H)$  so that

$$U \phi(f) U^{-1} = \phi(Af) \quad \forall f \in H?$$

(2) Given  $A$  as in (1) when is there an  $F$  in  $\Gamma(H)$  so that each  $\phi(f)$  is a Gaussian random variable with respect to  $|F|^2 d\mu$  but with variance  $\frac{1}{2} \|Af\|^2$  instead of  $\frac{1}{2} \|f\|^2$ .

(3) Given two Gaussian random fields with general covariance on the same Hilbert space  $H$  when can they be realized on a single measure space but with two mutually absolutely continuous measures?

The above problems are essentially equivalent and we present an answer to (3) under the assumption that  $A$  is positive and has a bounded inverse.

##### Theorem 4.4.1 (Feldman, Segal, Shale)

Let  $\phi : H \rightarrow L^2(\Omega, A, \mu)$  be a Gaussian unit random field and let  $A : H \rightarrow H$  be linear, bounded, positive with bounded inverse. Then a necessary and sufficient condition for the Gaussian random field on  $H$  with variance  $\frac{1}{2} \|Af\|^2$  to be realizable on  $\Omega$  with measure  $\nu$  equivalent to  $\mu$  is that  $A - I$  be a Hilbert Schmidt Operator. ■

#### 4.5 Discussion

Fock space has an important role to play in the study of the free Quantum field which can be considered to be an infinite assembly of non-interacting harmonic oscillators.

Consider the harmonic oscillator hamiltonian  $H = -\frac{1}{2} \frac{d^2}{dx^2} + \frac{1}{2} x^2$  as a self-adjoint operator on  $L^2(\mathbb{R})$ . If instead we work with  $L^2(\mathbb{R}, g)$ , where  $g$  is Gaussian measure then the action of  $H$  on the Hermite polynomials  $h_n$  is given by  $h_n \mapsto n h_n$ .

From this it is clear that the one-parameter  $\text{group } e^{itH}$  generated by  $H$  acts as  $h_n \mapsto e^{int} h_n$ . If we denote by  $W$  the unitary map  $f(x) \mapsto f(x) \left[ \frac{1}{\sqrt{2\pi}} e^{-x^2/2} \right]^{-1/2}$  :

$L^2(\mathbb{R}) \rightarrow L^2(\mathbb{R})$  then the operators  $p = i \frac{d}{dx}$  and  $q = x$  on  $L^2(\mathbb{R})$  transform to  $p' = WpW^{-1}$  and  $q' = WqW^{-1}$  and  $\frac{1}{2}(p'^2 + q'^2)$  can be identified as the Harmonic Oscillator hamiltonian. By the ideas of second quantization described in section 4.3, we can extend the one-dimensional case to an infinite dimensional setting. When this is done this gives rise to the particle representation of the free field and the concept of a number operator.

Shale's theorem on the equivalence of Gaussian Random fields essentially shows that the Kalman filter is analogous to the free quantum field. This can be done by noting that the observations and innovations processes are related by  $v = (I-K)y$  where  $K$  is a Hilbert-Schmidt operator.



Notes and References for Section 4

- (i) The basic reference for multiple Ito integrals is: K. Ito, Multiple Wiener Integrals, J. Math. Soc. Japan, 13, 1961, pp. 157-169.
- (ii) References for Section 4.1 are: D. Allinger and S.K. Mitter: New Results on the Innovations Problem, Technical Report R-964, Laboratory for Information and Decision Systems, M.I.T., January 1980. S.K. Mitter and D. Ocone: Multiple Integral Expansions for Non-linear Filtering, presented at the 18th IEEE CDC, Ft. Lauderdale, Florida, December 1979. D. Ocone, forthcoming Ph.D. Thesis, M.I.T. June 1980. S.I. Marcus, S.K. Mitter and D. Ocone, Finite Dimensional Non-linear Estimation for a class of Continuous and Discrete Time Problems, Proceedings of the International Conference on Analysis and Optimization of Stochastic Systems, Academic Press, 1980. For recent results, see E. Wong: Explicit Solutions to a Class of Non-linear Filtering Problems, to appear.
- (iii) The exposition on Random fields closely follows: B. Simon: The  $P(\phi_2)$ -Euclidean (Quantum) Field Theory, Princeton University Press, Princeton, N.J., 1975.
- (iv) Second Quantization as presented here is due to E. Nelson: The Free Markov Field, J. Functional Analysis, 12, 1973, pp. 211-227 where the basic Hypercontractivity Theorem is proved.
- (v) The idea of studying the infinite dimensional harmonic oscillator on Fock space is due to I.E. Segal: Tensor Algebras over Hilbert Spaces, I, Trans. Am. Math. Soc. 81, 1956, pp. 106-134.
- (vi) Viewing the Kalman Filter as the Free Quantum Field is due to the author: S.K. Mitter: Filtering Theory and Quantum Fields, to appear in Asterisque, 1980.

### 5. Some Lie Algebras of Interest in Filtering Theory.

Let us recall that in section 3 we showed that the unnormalized conditional density of  $x_t$  given the observations  $\{y_s | 0 \leq s \leq t\}$  satisfies a stochastic partial differential equation

$$(5.1) \quad dq(t, x, \omega) = L_0^* q(t, x, \omega) + L_1 q(t, x, \omega) dy_t$$

where  $L_0^*$  is the formal adjoint of the generator of the diffusion process corresponding to the stochastic differential

$$(5.2) \quad dx_t = f(x_t)dt + g(x_t)dw_t,$$

and hence  $L_0^*$  is given by

$$(5.3) \quad L_0^* = \frac{1}{2} \frac{d^2}{dx^2} g^2 - \frac{d}{dx} f,$$

and

$$(5.4) \quad L_1 = \text{Multiplication by } h(x).$$

The Lie Algebra of operators  $LA\{L_0^* - \frac{1}{2} L_1^2, L_1\}$  generated by  $L_0^* - \frac{1}{2} L_1^2$  and  $L_1$  have an important role to play in filtering theory. This comes about by writing equation (5.2) in Stratanovich form (a fact which can be justified rigorously under the hypotheses we have adopted)

$$(5.1)' \quad \frac{\partial q}{\partial t}(t, x) = (L_0^* - \frac{1}{2} L_1^2 + \dot{y}_t L_1) q(t, x).$$

This suggestive form of writing is meant to show the analogy of the filtering problem with the interaction picture of quantum physics.

It turns out that a number of filtering problems give rise to Lie Algebras of operators which are representations of known algebras which arise in mathematical physics. We first summarize these known facts about some Lie Algebras arising in

mathematical physics.

### 5.1. The Heisenberg Algebra and the Weyl Algebra.

Let  $n \geq 0$ , and let  $(p_1, \dots, p_n, q_1, \dots, q_n, z)$  be a basis for a real vector space  $V$ . On this space we can define the structure of a Lie Algebra by defining  $[p_i, q_i] = -[q_i, p_i] = z$ , the other brackets between elements of the basis being zero. Let us denote this Lie Algebra by  $N$ . The centre of  $N$  is  $\mathbb{R}z$  and we have  $[N, N] = \mathbb{R}z$  and hence this Lie algebra is nilpotent.

If  $N$  is a Lie algebra with centre  $c$  and if  $c = [N, N]$  and  $\dim c = 1$ , then  $N$  is a Heisenberg algebra.

#### Proposition 5.1

Let  $N$  be a nilpotent algebra. Suppose that all characteristic ideals<sup>(1)</sup> of  $N$  are of dimension  $\leq 1$ . Then either  $N$  is zero or a Heisenberg algebra.  $\square$

We denote the algebra defined by  $2n$ -generators  $p_1, q_1, \dots, p_n, q_n$  and the bracket relations

$$[p_i, q_i] = 1.$$

$$[p_i, q_j] = [p_i, p_j] = [q_i, q_j] = 0 \text{ for } i \neq j$$

by  $A_n(\mathbb{R})$  or simply  $A_n$ . These are the so-called Weyl algebras.

Elements of the form  $p_1^{i_1} q_1^{j_1} \dots p_n^{i_n} q_n^{j_n}$  ( $i_1, j_1, \dots, i_n, j_n \in \mathbb{N}$ ) generate the vector space  $A_n$ .

In the vector space  $E = \mathbb{R}[X_1, \dots, X_n]$  let  $P_i$  be the endomorphism  $\frac{\partial}{\partial X_i}$  and  $Q_i$  the endomorphism of multiplication by  $X_i$ . We have

$$[P_i, Q_i] = 1.$$

$$[P_i, Q_j] = [P_i, P_j] = [Q_i, Q_j] = 0 \text{ for } i \neq j$$

---

(1) Let  $g$  be a Lie algebra. A derivation of  $g$  is a linear mapping  $D: g \rightarrow g$  such that  $D([x, y]) = [Dx, y] + [x, Dy]$   $\forall x, y \in g$ . A characteristic ideal of  $g$  is a subspace which is stable for all derivations of  $g$ .

and hence there exists a homomorphism  $\rho: A_n \rightarrow \text{End}(E)$  such that  $\rho(p_i) = P_i$  and  $\rho(q_i) = Q_i$  for all  $i$ . The elements  $P_1^{i_1} Q_1^{j_1}, \dots, P_n^{i_n} Q_n^{j_n}$  are linearly independent and therefore the elements  $P_1^{i_1} Q_1^{j_1}, \dots, P_n^{i_n} Q_n^{j_n}$  form a basis for the vector space  $A_n$  and  $\rho$  is injective. It follows that  $A_n \cong A_1 \otimes \dots \otimes A_1$  ( $n$ -copies). The representation  $\rho$  of  $A_n$  in  $E$  is termed the standard representation. Finally  $E$  is a simple  $A_n$ -module and the set of  $A_n$ -endomorphisms of  $E$  is  $\mathbb{R}$ .

Let  $B_m$  be the set of linear combinations of

$$P_1^{i_1} Q_1^{j_1}, \dots, P_n^{i_n} Q_n^{j_n} \in A_n \text{ such that } i_1 + j_1 + \dots + i_n + j_n \leq m$$

Then  $B_m \subset B_{m+m'}$ .

Consequently the graded algebra associated with  $A_n$  equipped with the filtration  $(B_0, B_1, \dots)$  is the polynomial algebra in  $2n$ -variables.

#### Proposition 5.2

- (i)  $A_n$  is integral and noetherian.
- (ii) The centre of  $A_n$  is  $\mathbb{R}$ .
- (iii) The algebra  $A_n$  is simple.  $\blacksquare$

#### Proposition 5.3

Let  $p_1, q_1, \dots, p_n, q_n$  be the canonical generators of  $A_n$ . Let us define vector subspaces  $S, T$  of  $A_n$  as follows:

$$(i) \quad S = \sum_{i=1}^n \{ (rp_i + rq_i) \mid r \in \mathbb{R} \}.$$

$$(ii) \quad T = \sum_{1 \leq i, j \leq n} \left\{ \frac{r}{2} (p_i q_j + q_j p_i) + rp_i p_j + rq_i q_j \mid r \in \mathbb{R} \right\}.$$

Then:

- (i)  $\mathbb{R} \oplus S$  and  $T$  are Lie sub-algebras of  $A_n$
- (ii)  $\mathbb{R} \oplus S$  is a Heisenberg algebra.  $\blacksquare$

The relationship between Heisenberg and Weyl algebras is given by

Proposition 5.4

$$A_n(\mathbb{R}) \cong U(N)/J$$

where  $U(N)$  is the enveloping algebra of the  $(2n+1)$ -dimensional Heisenberg Algebra  $N$  and  $J$  is a two-sided ideal in  $U(N)$ .  $\blacksquare$

5.2. The Oscillator Algebra.

The real Lie algebra with  $2n + 2$  generators  $(h, p_1, \dots, p_n, q_1, \dots, q_n, z)$  satisfying the bracket relations

$$[h, p_i] = q_i, \quad [h, q_i] = p_i, \quad [p_i, q_i] = z, \text{ and the other brackets being zero}$$

is defined to be the oscillator algebra. This is a solvable Lie algebra. This algebra is the semi-direct sum of the one-dimensional algebra spanned by  $h$  and the Heisenberg algebra  $N$ .

5.3. The Poisson Bracket Algebra.

Consider  $\mathbb{R}^{2n} = \{(q, p) \mid q = (q_1, \dots, q_n), p = (p_1, \dots, p_n)\}$  with its standard symplectic structure (see later section). For  $f, g \in C^\infty(\mathbb{R}^{2n})$ , the Poisson bracket is given by

$$\{f, g\} = \sum_{i=1}^n \left( \frac{\partial f}{\partial q_i} \cdot \frac{\partial g}{\partial p_i} - \frac{\partial f}{\partial p_i} \cdot \frac{\partial g}{\partial q_i} \right)$$

The real vector space of polynomials in  $(q, p)$  denoted by  $\mathcal{P}_n$  is a Lie algebra under the Poisson bracket operation.

Proposition 5.5

$\mathcal{P}_n$  is generated by the two polynomials  $q_1$  and  $h = p^2 + q_1^4$  if  $n = 1$  and  $q_1$  and  $h = \sum_{i=1}^n (p_i^2 + q_i^4) + \sum_{i=1}^{n-1} q_i q_{i+1}$  if  $n > 1$ . That is  $\mathcal{P}_n$  coincides with the smallest Lie subalgebra  $\mathcal{Q}_n$  containing  $q_1$  and  $h$ .

Proposition 5.6

$P_n/\mathbb{R}$  is simple.  $\blacksquare$

5.4 Examples of Non-linear Filtering Problems and Their Lie Algebras

Example 1.

Consider the non-linear filtering problem

$$(5.5) \quad \begin{cases} dx_t = f(x_t)dt + dw_t \\ dy_t = h(x_t)dt + dn_t \end{cases}$$

and let us assume that  $f, h \in C^\infty(\mathbb{R})$ . In the light of section 3, the Mortensen-Zakai equation is

$$dq_t = (L_0^* - \frac{1}{2} L_1^2)q_t dt + L_1 q_t \cdot dy_t$$

where  $L_0^* = \frac{1}{2} \frac{d^2}{dx^2} - \frac{d}{dx} f$

$L_1$  = multiplication by  $h(x)$ .

We then have

Proposition 5.7

$LA\{L_0^* - \frac{1}{2} L_1^2, L_1\}$  is finite dimensional only in the case

(i)  $h = \alpha x + \beta$ .

(ii)  $f_x + f^2 = ax^2 + bx + c$  ,

where the Lie algebra of operators is computed on the common domain  $C^\infty(\mathbb{R})$  or  $S(\mathbb{R})$ .  $\blacksquare$

In the above if we want the diffusion process to be defined globally on  $\mathbb{R}$ , so we assume that the Riccati equation  $f_x + f^2 = ax^2 + bx + c$  has a global solution.

The example covered by Proposition 5.7 has recently been considered by Benes. From Section 3.1, we can see that we can remove the effect of the drift by a Gauge

transformation. From (3.15) we see that it is enough to solve

$$d\tilde{q}_t = \hat{L}_0 \tilde{q}_t dt - V(x) \tilde{q}_t dt + h \tilde{q}_t \cdot dy_t$$

But  $LA\{\hat{L}_0^* - \frac{1}{2}L_1^2, L_1\} \approx LA\{\hat{L}_0 - V(x), L_1\}$ , and under the hypotheses of Proposition 5.7 the second Lie algebra has a basis consisting of the elements

$$\hat{L}_0 - \frac{1}{2}L_1^2, L_1, [\hat{L}_0 - \frac{1}{2}L_1^2, L_1] = L_2 = \alpha \frac{d}{dx}, \quad [L_2, L_2] = \alpha I.$$

This however is essentially the Harmonic Oscillator algebra and corresponds (essentially) to the filtering problem

$$\begin{cases} x_t = w_t \\ dy_t = x_t dt + d\eta_t \end{cases}$$

which is a Kalman filtering problem.

It should also be remarked that there is no difficulty in extending this example to the multi-dimensional situation

$$dx_t = f(x_t)dt + dw_t$$

$$dy_t = Hx_t dt + d\eta_t,$$

provided  $f: \mathbb{R}^n \rightarrow \mathbb{R}^n$  is  $\nabla F = (\frac{\partial F}{\partial x_1}, \dots, \frac{\partial F}{\partial x_n})'$  and  $H: \mathbb{R}^n \rightarrow \mathbb{R}^p$ . This fact is also clear from our considerations relating Schrödinger and Dirichlet operators. Finally, this problem is the analogue of the imaginary time Harmonic Oscillator problem with an external force.

Example 2.

Consider the non-linear filtering problem

$$(5.6) \quad \begin{aligned} x_t &= w_t \\ dy_t &= x_t^3 dt + d\eta_t \end{aligned}$$

This is the so-called cubic sensor problem. From Proposition 5.7,  $LA\{\frac{1}{2}\frac{d^2}{dx^2} - \frac{1}{2}x^6, x^3\}$  is infinite dimensional. From the calculations required to establish Proposition 5.7 it can be shown that the Lie algebra is isomorphic to the Weyl algebra  $A_1$  which is simple.

Example 3.

Consider the model given by (5.5) and suppose that  $f_x + f^2 = V(x)$  has a global solution where  $V(x)$  say is an even positive polynomial (other than the quadratic). Then from Proposition 5.7,  $LA\{L_0^* - \frac{1}{2}L_1^2, L_1\}$  is infinite dimensional. Even more, these Lie algebras (modulo their centre which is  $R$ ) are simple. This fact follows by doing calculations similar to those involved in establishing that the Poisson bracket algebra is simple.

Example 4.

Let us consider the model of (5.5) and let  $f, h \in C^\infty(U)$  where  $U$  is some open set in  $R$ . Then by restricting  $L_0^* - \frac{1}{2}L_1^2$  and  $L_1$  to  $C^\infty(U)$  the Lie algebra  $LA\{L_0^* - \frac{1}{2}L_1^2, L_1\}$  is finite dimensional only in the (prototype) case

- (i)  $h = x^2$
- (ii)  $f_x + f^2 = -\frac{1}{4}x^4$



Note that the last equation (Riccati) has finite escape time.

Suppose we consider the slightly more general model

$$dx_t = f(x_t)dt + g(x_t)dw_t, \text{ where } f, g \in C^\infty(U)$$

$U$  open in  $R$  and  $g(x) > 0$ ,  $x \in U$ . In this case we can remove the diffusion term  $g$  by a non-linear smooth change of coordinates.

Let  $x_t = \alpha(z_t)$  where  $\alpha$  is smooth and invertible. Then by the Ito differential rule

$$dx_t = \alpha_z dz_t + \frac{1}{2} \alpha_{zz} dt = f(\alpha(z_t))dt + g(\alpha(z_t))dw_t.$$

Hence

$$dz_t = (\alpha_z)^{-1}(f(\alpha(z_t)))dt - \frac{1}{2} \alpha_{zz}(z_t)dt + (\alpha_z)^{-1}g(\alpha(z_t))dw_t$$

Let  $\alpha$  satisfy the differential equation

$$\frac{d\alpha}{dz} = g(\alpha(z)),$$

and let  $I$  be the maximal interval in  $R$  on which the solution is defined. Then

$$dz_t = (g(\alpha(z_t)))^{-1}f(\alpha(z_t)) - \frac{1}{2}g_\alpha(z_t)dt + dw_t = \tilde{f}(z_t)dt + dw_t.$$

The observation equation is

$$dy_t = \tilde{h}(z_t)dt + d\eta_t \text{ where } \tilde{h} = h(\alpha(z_t)).$$

Now it is easy to see that

$$LA\left\{\frac{1}{2} \frac{d^2}{dx^2} g^2 - \frac{1}{2} \frac{d}{dx} f - \frac{1}{2} h^2, h\right\} = LA\left\{\frac{1}{2} \frac{d^2}{dx^2} - \frac{d}{dx} \tilde{f} - \frac{1}{2} \tilde{h}^2, \tilde{h}\right\}$$

where the first Lie algebra is computed on the common invariant domain  $C^\infty(U)$  and the second on the common invariant domain  $C^\infty(I)$ .

### 5.5 Significance of the Results

If the Lie Algebra  $LA\{L_0^* - \frac{1}{2} L_1^2, L_1\}$  is finite dimensional, then it should be possible to construct the filter by integrating the Lie algebra. We consider this aspect of the problem in the next section. If the Lie algebra is infinite dimensional and simple then we conjecture that it cannot be represented by a Lie algebra of vector fields with analytic coefficients. That is, we cannot represent the solution  $q(t, z, y_0^t)$  of the Mortensen-Zakai equation by means of a finite dimensional sufficient statistic described by a vector differential equation

$$(5.7) \quad d\alpha_t = a(\alpha_t)dt + b(\alpha_t)dy_t .$$

Suppose we are interested only in computing unnormalized conditional statistics  $\hat{\phi}(x_t)$ . If we denote this by  $\hat{\phi}_t$ , then

$$\hat{\phi}_t = \int_{\mathbb{R}} \phi(z) q(t, z, y_0^t) dz .$$

It may be possible to represent  $\hat{\phi}_t$  as the output of a vector differential system:

$$(5.8) \quad \begin{cases} d\alpha_t = a(\alpha_t)dt + b(\alpha_t)dy_t \\ \phi_t = c(\alpha_t) \end{cases}$$

A theorem of Brockett says that for this to be possible there must exist a homomorphism between the Lie algebra of operators  $LA\{L_0^* - \frac{1}{2} L_1^2, L_1\}$  and the Lie algebra of vector fields  $a - \frac{1}{2} b_x b$  and  $b$  such that under the homomorphism

$L_0^* = \frac{1}{2} L_1^2 \mapsto a - \frac{1}{2} b_x b$  and  $L_1 \mapsto b$ . This would suggest that the ideal structure of the Lie Algebra  $L = LA\{L_0^* = \frac{1}{2} L_1^2, L_1\}$  should be important. If there exist non-trivial ideals  $I$  then a candidate for such a homomorphism would be  $\psi : L \rightarrow L/I$ . The results of this section show that for a large class of problems this line of attack will not be possible, since the Lie algebra  $L$  is infinite dimensional and simple.

A possible approach might be to represent  $\hat{\phi}_t$  as the output of a delay system, that is look for a representation of the form

$$\begin{cases} d\alpha_t = a(\alpha_t, \alpha_{t-\theta})dt + b(\alpha_t, \alpha_{t-\theta})dy_t \\ \hat{\phi}_t = \int_{-\theta}^0 c(\alpha_t, \alpha_{t-\theta})d\theta \end{cases}$$

There is some reason to believe that it might be possible to do this for the cubic sensor problem considered in Example 2.

Notes and References for Section 5

(i) The idea of studying the Lie Algebra of operators involved in the Mortensen-Zakai equation is independently due to R.W. Brockett and myself. I was motivated by the analogy between the Kalman filter and the Free Quantum field and I discussed these ideas in a series of lectures in the University of Maryland in December 1977.

In <sup>this</sup> connection see: R.W. Brockett: Remarks on Finite Dimensional Non-linear Estimation, presented in the Conference on Algebraic and Geometric Methods in System Theory, Bordeaux, France, September 1978, to appear in Asterisque, 1980.

S.K. Mitter: Filtering Theory and Quantum Fields, presented in the Conference on Algebraic and Geometric Methods in System Theory, Bordeaux, France, September 1978, to appear in Asterisque, 1980 and the references cited in the above two papers.

(ii) For the material on Heisenberg and Weyl Algebras see: J. Dixmier: Algèbres Enveloppantes, Gauthier-Villars, Paris, 1974.

Propositions 5.5 and 5.6 on the Poisson Bracket Algebra are due to A. Avez and A. Heslot: L'algebre de Lie des Polynômes en les Coordonnées Canoniques munie de Crochet de Poisson, C.R. Acad. Sc. Paris t. 288 Serie A, Mai 1979, pp. 831-833.

(iii) Proposition 5.7 is due to Daniel Ocone and is related to non-linear filtering problems first introduced by V. Benes. See D. Ocone: Doctoral Dissertation, Mathematics Department, M.I.T., June 1980 and V. Benes: paper to appear in Stochastics, 1980.

(iv) The content of Examples 2 and Examples 3 are new and due to Daniel Ocone and the author. The content of Example 4 is due to Daniel Ocone (forthcoming Ph.D. dissertation) and R.W. Brockett: loc. cit, paper in Decision and Control Conference 1979.

(v) The reference to the theorem of Brockett cited in section 5.5 is contained in R.W. Brockett: loc. cit.

(vi) Recent work (unpublished) by John Baras as well as M. Hazewinkel and S. Marcus shows that the bilinear filtering has a nice ideal structure and many filterations.

## 6. Representation of the Filter

The complete solution of the non-linear filtering problem requires computing the propagator of the Mortensen-Zakai equation or equivalently the unnormalized conditional transition density  $q(t, z, y_0^t | s, x)$ . We have discussed several approaches to computing the density  $q(t, z, y_0^t)$ . For the class of problems considered by Benes the propagator could be computed by solving the two-point boundary value problem

$$d\xi_\tau = -\eta_\tau d\tau; \xi_s = x, \xi_t = z$$

$$d\eta_\tau = -\xi_\tau d\tau - V_\xi(\xi_\tau) d\tau + dy_\tau$$

according to the development of section 3.2. This is the approach of functional integration.

The Lie-algebra viewpoint to computing the propagator would require that the Lie algebra be integrable. This is a difficult question as we can see by discussing the commutation relationship of quantum mechanics.

### 6.1 Canonical Commutation Relations and their Unitary Representations.

Consider a massive spinless, nonrelativistic particle. Its configuration space is typically  $L^2(\mathbb{R})$ . The position operator  $q : L^2(\mathbb{R}) \rightarrow L^2(\mathbb{R})$  is defined

as the operator with domain

$$(6.1) \quad \begin{cases} \mathcal{D}(q) = \{f \in L^2(\mathbb{R}) \mid xf(x) \in L^2(\mathbb{R})\} & \text{and} \\ (qf)(x) = x f(x). \end{cases}$$

The momentum operator  $p : L^2(\mathbb{R}) \rightarrow L^2(\mathbb{R})$  has domain

$$(6.2) \quad \begin{cases} \mathcal{D}(p) = \{f \in L^2(\mathbb{R}) \mid f' \in L^2(\mathbb{R})\} & \text{and} \\ pf(x) = i \frac{df}{dx}. \end{cases}$$

The Schwartz space  $S(\mathbb{R}) \subset \mathcal{D}(q) \cap \mathcal{D}(p)$ ,  $S(\mathbb{R})$  is dense in  $L^2(\mathbb{R})$  and is left invariant under  $q$  and  $p$ . Moreover on  $S(\mathbb{R})$  they satisfy the Heisenberg Commutation Relations

$$(6.3) \quad qp - pq = i I$$

Since  $p$  is self-adjoint, the operator  $U(a) = e^{-ia p}$ ,  $a \in \mathbb{R}$  is unitary and the operators

$\{U(a) \mid a \in \mathbb{R}\}$  forms a one-parameter strongly continuous unitary group.

Now  $\mathcal{D}(q)$  is invariant under  $U(a)$ ,  $a \in \mathbb{R}$  and it can be shown that

$$(6.4) \quad U(a)qU(-a) = q - a I \text{ on } \mathcal{D}(q).$$

This is the Schrödinger form of the Canonical Commutation Relations.

Since  $q$  is self-adjoint it generates a strongly continuous one-parameter unitary group  $V(b)$  given by

$$(6.5) \quad V(b)f(x) = e^{-ibx}f(x).$$

It can be shown that

$$(6.6) \quad U(a)V(b) = e^{iab}V(b)U(a)$$

This is done by first checking it on  $S(\mathbb{R})$  and then extending by continuity to all of  $L^2(\mathbb{R})$ .

(6.6) represents the Weyl Form of the Commutation Relations.

It is a well known fact that the Heisenberg form of the canonical commutation relations do not give rise to a unitary group representation, while the Schrödinger and Weyl forms do. This gives rise to the following general question: Suppose we have a representation of a Lie algebra  $\mathfrak{g}$  by skew-symmetric operators defined on a common invariant domain  $\mathcal{D}$  in a Hilbert space  $H$  and let  $G$  be the simply connected Lie group with Lie algebra  $\mathfrak{g}$ . When does the representation of  $\mathfrak{g}$  come from a unitary representation of  $G$ ?

The answer to this question is connected with questions of essential self-adjointness and the existence of a common dense set of analytic vectors. We discuss some of these questions now.

Let  $G$  be a Lie group and  $X$  a Banach space. A representation  $T$  of  $G$  is a mapping  $G \rightarrow L(X) : \alpha \mapsto T(\alpha)$ , where  $L(X)$  is the set of bounded operators such that  $T(e) = I$ ,  $e$  being the identity element and  $T(\alpha_1 \cdot \alpha_2) = T(\alpha_1)T(\alpha_2)$ , for all  $\alpha_1, \alpha_2 \in G$  and  $\forall x \in X$ ,  $\alpha \mapsto T(\alpha)x$  is continuous (with  $X$  the norm topology). The representation is called unitary if  $X$  is a Hilbert space and each  $T(\alpha)$  is a unitary operator.

A vector  $x \in X$  is an analytic vector for  $T$  in case the mapping  $\alpha \mapsto T(\alpha)x : G \rightarrow X$  is analytic.

The salient facts connecting representations and analytic vectors are:

Theorem 6.1: Let  $T$  be a representation of a Lie group  $G$  on a Banach space  $X$ . Then  $T$  has a dense set of analytic vectors in  $X$ .  $\blacksquare$

The answer to the question raised earlier in the section is contained in the following theorem and corollary due to Nelson:

Theorem 6.2: Let  $\mathfrak{g}$  be a Lie algebra of skew symmetric operators on a Hilbert space  $H$  having a common invariant domain  $\mathcal{D}$ . Let  $X_1, \dots, X_d$  be a basis for  $\mathfrak{g}$ ,  $\Delta = X_1^2 + \dots + X_d^2$ . If  $\Delta$  is essentially self-adjoint, then there is on  $H$  a unique unitary representation  $U$  of the simply connected Lie Group  $G$  having  $\mathfrak{g}$  as its Lie algebra such that for all  $X$  in  $\mathfrak{g}$ ,  $\overline{U(X)} = \bar{X}$  (bar denotes closure of an operator).  $\blacksquare$

Corollary 6.3: Let  $g$  be a real Lie Algebra,  $H$  a Hilbert space. For each  $X$  in  $g$  let  $\rho(X)$  be a skew-symmetric operator on  $H$ . Let  $\mathcal{D}$  be a dense linear subspace of  $H$  such that for all  $X, Y$  in  $g$ ,  $\mathcal{D}$  is contained in the domain of  $\rho(X)\rho(Y)$ . Suppose that for all  $X, Y$  in  $g$ ,  $x$  in  $\mathcal{D}$ , and real numbers  $a$  and  $b$ ,

$$\begin{aligned}\rho(aX + bY)x &= a\rho(X)x + b\rho(Y)x \\ \rho([X, Y])x &= (\rho(X)\rho(Y) - \rho(Y)\rho(X))x\end{aligned}$$

Let  $X_1, \dots, X_d$  be a basis for  $g$ . If the restriction  $A$  of  $\rho(X_1)^2 + \dots + \rho(X_d)^2$  to  $\mathcal{D}$  is essentially self-adjoint, then there is on  $H$  a unique unitary representation  $U$  of the simply connected Lie group  $G$  having  $g$  as its Lie algebra such for all  $X$  in  $g$ ,  $\overline{U(X)} = \overline{\rho(X)}$ .

Corollary 6.4: Let  $g$  be a real Lie algebra with a basis  $X_1, \dots, X_d$ ,  $G$  the simply connected Lie group with Lie algebra  $g$ ,  $H$  a Hilbert space,  $\mathcal{C}$  a dense linear subspace of  $H$ . Let  $\rho$  be a representation of  $g$  by skew-symmetric operators with domain  $\mathcal{C}$ . Then there is a unitary representation  $U$  of  $G$  such that  $\mathcal{C}$  is the space of infinitely differentiable vectors for  $U$  and  $U(X) = \rho(X)$  for all  $X$  in  $g$  if and only if

$$A = \rho(X_1)^2 + \dots + \rho(X_d)^2$$

is essentially self-adjoint and  $\mathcal{C} = \bigcap_{n=1}^{\infty} \mathcal{D}(\bar{A}^n)$ .

How do these ideas relate to the representation of the filter? Firstly, the equation we are dealing with is a stochastic parabolic equation valid for  $t \geq 0$ . Hence the operators  $L_0^*$  and  $L_1$  will in general only generate semi-groups. Consider the Kalman filtering problem (or any problem Gauge equivalent to it). Then what is necessary is to give a precise meaning to the time-ordered operator product

$$e^{t_1 L_0^*} e^{t_2 L_1} e^{t_3 L_2} e^{t_4 L_3} \quad , \quad t_i \geq 0$$

as an evolution operator (in the sense of Kato). For the Kalman filtering this can



be done using special methods. There are examples, like the estimation of a Bessel process in additive white noise where this appears not possible to do. To see the connection to unitary representations of Lie groups, it might be best to complexify and try to check the conditions of Nelson's theorem. We conjecture that if the Lie algebra representation does not extend to a unitary group representation then we shall not be able to give meaning to the time-ordered operator product considered above.

Finally, the most direct way appears to be to try to integrate equation (3.15) of Section 3

$$\frac{d\tilde{q}_t}{dt} = (L_0 - \frac{1}{2} L_1^2)\tilde{q}_t + y_t L_2 \dot{\tilde{q}}_t - y_t^2 L_3 \tilde{q}_t$$

but the question of integrating the Lie algebra also appears here.

Notes and References for Section 6

(i) For a detailed account of group representation theory relevant to this section see, A. O. Barut and R. Raczk: Theory of Group Representations, PWN-Polish Scientific Publishers, Warsaw 1977. The exposition given here closely follows:

E. Nelson, Analytic Vectors, *Annals of Mathematics*, 70, 1959, pp. 572-615.

(ii) The details of treating the Benes and similar problems using Group Invariance and Lie Algebraic ideas will appear in a joint paper by J. Baras, S.K. Mitter, D. Ocone. See also D. Ocone, forthcoming Doctoral Dissertation, M.I.T., June 1980.

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